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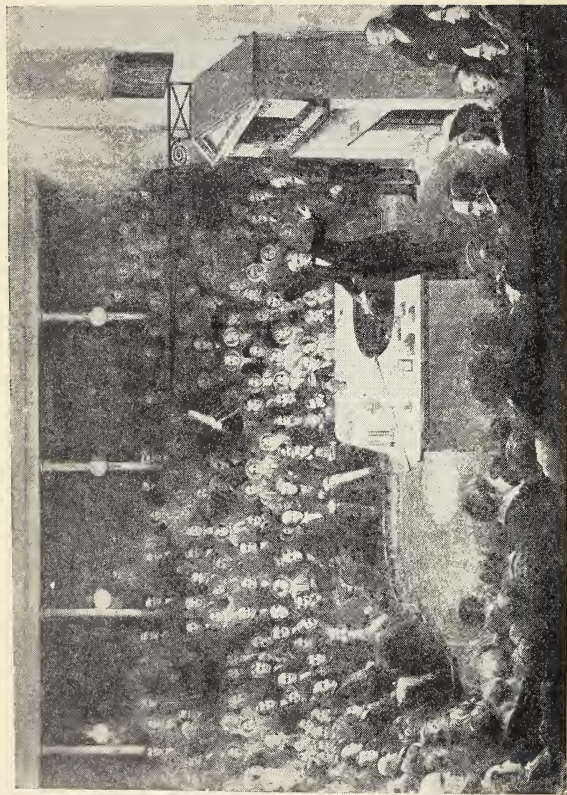
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ELEMENTARY PHYSICS



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ELEMENTARY PHYSICS

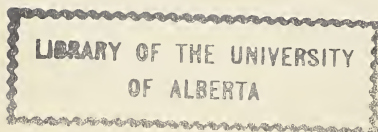
By

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PREFACE

As already explained in the preface to *A Junior Physics*, the present volume is the second and concluding part of a concentric course in Physics. The earlier volume is intended for pupils of eleven to fourteen or thereabouts, and it is hoped that its successor will be found suitable for the remaining period of a course which is normally completed at age sixteen.

While there is obviously the closest relation between the two books, I have tried to arrange that the present volume shall at the same time be reasonably independent of its predecessor. Accordingly, most of the chapters which follow will be found to include a brief recapitulation of the corresponding work already done, followed by a fuller treatment of the particular subject in hand. This fuller treatment often involves a greater emphasis on the quantitative and mathematical side. Thus the subjects of Expansion, Thermal Capacity, Latent Heat, Mechanical Equivalent, Refraction, Magnetic Fields, Electric Current, Electrolysis—and the list is by no means exhaustive—can very properly be presented in their simpler qualitative aspects to quite young pupils, but a fuller treatment is needed later, with greater emphasis on the quantitative side.

The book closes with a few pages on the subjects of radio-location and atomic energy. Obviously nothing more than the merest sketch could be attempted in a volume of this character, but it is good for a boy to realise that all the little fields he has entered in his school studies of Physics are as nothing compared with the vast territory, still largely unexplored, which lies beyond.

I would again like to record my deep indebtedness to Mr. D. H. Turner, B.Sc., formerly my colleague at Hele's School, Exeter, and to my son-in-law, Mr. R. C. Tudway, M.B., B.S., B.Sc., both of whom read through the manuscript and made many valuable suggestions. My son-in-law also gave me much help in reading the proofs. I must also thank the various

authorities concerned for their kind permission to include questions set in recent years at the School Certificate Examinations arranged by the Universities of Bristol, Cambridge, Durham, London and Oxford, by the Central Welsh Board, the Oxford and Cambridge Schools Examination Board, and the Joint Matriculation Board. Finally, I wish to thank a number of business firms and private individuals for help in securing suitable half-tone illustrations. Their courtesy has been acknowledged on the appropriate pages.

W. LITTLER

Bristol,
Sept. 1948

PREFACE TO SECOND EDITION

IN this edition there is a rather fuller treatment of alternating current, and explanations relating to electrolysis have been brought into line with modern theory. A few additional applications of principles already dealt with (e.g. the light meter in relation to photometry) have been given, and the underlying connection between heat, light, 'wireless' transmission and other phenomena has been discussed and illustrated. In general, the page numbering has been altered either very little or not at all, and there should be no serious difficulty in using this edition side by side with the earlier one.

W. LITTLER

Jan. 1955

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Questions

Many of the questions in this book are reproduced from School Certificate papers of the various examining bodies. To indicate the source the following contractions are used :—

Brist. = *Bristol.*

Camb. = *Cambridge.*

C.W.B. = *Welsh Joint Education Committee.*

Dur. = *Durham.*

Lond. = *London.*

J.M.B. = *Joint Matriculation Board.*

O.C. = *Oxford and Cambridge Schools Examination Board.*

Oxf. = *Oxford.*

The addition of an asterisk (e.g. *Lond.**) means that only part of the original question has been used.

Often the questions are divided into two groups, A and B. The 'A' questions are the easier ones.

PART I

MECHANICS AND HYDROSTATICS

CHAPTER 1

VELOCITY

FOR at least two thousand years men have interested themselves in questions connected with motion. Aristotle, for instance, made some study of falling bodies, and came to the conclusion (a wrong one as it turned out) that the rate at which a body falls is proportional to its weight. Thus, according to him, if a 10-lb. weight and a 1-lb. weight were dropped, the first weight would fall ten times as fast as the second. Galileo (1564-1642) first cleared the ground for further progress by proving that Aristotle's conclusion was false, and then set himself to unravel the whole problem of what we may call the Law of Falling Bodies. He was greatly handicapped by the lack of suitable apparatus, but in spite of this, by making the best use of what he had, he made marvellous progress. Sir Isaac Newton added results and conclusions of his own to those obtained by Galileo, and developed the science of Mechanics (which includes the study of Motion) practically to the form in which we have it to-day.

Let us begin by thinking of the terms *speed* and *velocity*. *Speed*, we may say, is *the rate at which a body is changing its position*. The question of *direction* does not enter into the matter. *Velocity*, on the other hand, is *the rate at which a body is changing its position in a given direction*.

We can illustrate the difference between the two by thinking of a toy railway train running round a circular track, at a steady rate of, say, 2 ft. per sec. Its speed is constant, but its velocity is constantly changing because its direction is changing. Strictly speaking, we have not completely stated the velocity of a body unless we have mentioned *both* its speed

and the direction in which it is moving. All the same, expressions such as 'a velocity of 10 ft. per sec.' are often met with. In that case it is implied that this movement is taking place in a constant direction.

Representation. Velocity is conveniently represented by a straight line, the length of the line standing for the magnitude of the velocity, and its direction indicating the direction of the velocity. The 'sense' of the movement (e.g. whether from east to west or west to east) is usually shown by an arrow.

Fig 1/1 two velocities are represented, one of $7\frac{1}{2}$ m.p.h. to the N.E., and another of 5 m.p.h. to the S.E. If OP measured 1 in., the scale would be 1 in. = 5 m.p.h.

Uniform velocity. A body is said to move with **uniform velocity** when it passes over equal distances in equal intervals of time, however short these intervals may be, the direction remaining constant.

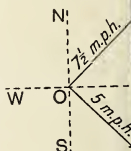


FIG. 1/1

A motor car travelling north may have completed 30 miles at the end of the first hour, 60 at the end of the second, 90 at the end of the third and so on, yet it is not practically certain that its velocity has not been constant. It will have had to slow down when passing through towns, making up for lost time when it reaches an open road in the country. All we can say is that its *average* velocity is 30 miles an hour.

If, however, the car passes successive milestones at intervals of exactly 2 mins., it is getting considerably nearer to our standard of 'uniform velocity.' If, second after second, it covers exactly 44 ft. (30 miles per hour = 44 ft. per sec.) we might then say that its velocity is practically uniform.

Units employed. In the f.p.s. system of units (foot-pound-second) velocities are expressed in feet per second, while in the c.g.s. system they are expressed in centimetres per second. The usual 'everyday-life' units, however, are miles per hour and (on the Continent) kilometres per hour.

It is very useful to remember that 60 miles per hour = 88 feet per second. Remembering this, we can very quickly reduce such velocities as 30 miles per hour, 10 miles per hour etc., to feet per second.

Space-time diagram. Suppose a man is walking with

uniform velocity of 6 ft. per sec. (= 4 m.p.h. approximately). Let us express his movement graphically.

We draw two lines at right angles, OX horizontally, and OY vertically. Along OX we indicate time to a scale of, say, 1 in. = 1 sec. Along OY we represent distance or 'space.' In this case a convenient scale would be 1 in. to represent 6 ft.

O represents the point at which we begin to record the man's movement (space and time both zero). After 1 sec. he has covered 6 ft., and we get the point A. Similarly we obtain B (2 sec. 12 ft.), C, D, etc. By joining all these points we have what is called a 'space-time' or 'displacement-time' curve (or diagram). It can easily be proved that where the velocity is uniform, as in this case, the space-time diagram consists of a straight line.

Notice that the velocity, being $\text{space} \div \text{time}$, is represented by the quotient $DK \div OK$, a quotient known as the *slope* or *gradient* of the curve.

Obviously we should obtain the same answer if we took the quotient $CL \div OL$, $CM \div AM$,

or $HN \div PN$, etc. That is because, in the case under consideration, the velocity is uniform.

Variable velocity. Let us now consider the space-time curve for a body travelling with variable velocity. Suppose, for instance, a car starts from rest, and lines are chalked across the road at intervals of 25 ft. It crosses these lines after the following intervals, expressed in seconds: 2.5, 4.2, 5.4, 6.6, 7.6, 8.4, 9.2, 10.0. Let us represent a second by half an inch horizontally, and 25 ft. by half an inch vertically. Here is our curve (Fig. 1/3).

Take the last point on the curve (D), and draw the perpendicular DK. DK gives the distance (200 ft.) travelled in the time represented by OK (10 sec.). Thus $DK \div OK$ or $200 \text{ ft.} \div 10 \text{ sec.} = 20 \text{ ft. per sec.}$ is the *average* velocity for the whole period.

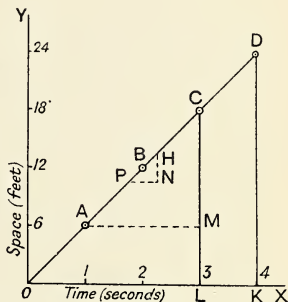


FIG. 1/2

Let us next consider the velocity at a particular instant, say that represented by A, 4.2 sec. after the start. Now, strictly speaking, a velocity can only be recorded over an *interval* of time, and as an 'instant' is not an interval, it looks at first sight as though we cannot find the velocity under discussion. However, if we consider an *extremely short* interval of time beginning just before the instant A and ending just after it, the velocity during that interval will be for practical purposes the velocity we require, for in such a short interval it will have had no chance to vary appreciably.

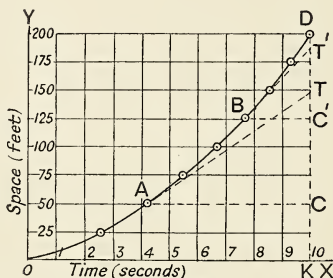


FIG. 1/3

So we may go back to our question—to find the velocity at the instant represented by A. In the space-time diagram on p. 3 we saw that the velocity was given by the *slope* of the curve. Similarly, when the velocity is variable, the velocity at any instant is given by the slope of the curve *at the point representing that instant*.

Now, the slope of a curve at a particular point is represented by the *tangent* to the curve at that point, so at length we see our way to a solution. Using a ruler as carefully as possible draw a tangent at A, cutting DK in T. Draw AC parallel to the time-axis, cutting DK in C. Measure TC (= 100 ft.) and AC (= 5.8 sec.). Required velocity in ft. per sec. = $100 \div 5.8 = 17.2$. Similarly velocity at instant B (i.e. 7.6 sec. from the start) = $C'T' \div BC' = 65 \text{ ft.} \div 2.4 \text{ sec.} = 21.1 \text{ ft. per sec.}$

Example 1. A body is projected vertically upwards, and is found to be 36, 64, 84, 96, 100, 96, 84, 64 and 36 ft. from the ground after $\frac{1}{2}$, 1, $1\frac{1}{2}$, 2, $2\frac{1}{2}$, 3, $3\frac{1}{2}$, 4 and $4\frac{1}{2}$ sec., respectively. Construct the space-time curve for the motion. Explain how the velocity of the body at any instant can be deduced from the diagram, and in this way deduce the velocity of the body (a) $1\frac{1}{2}$ sec., (b) $2\frac{1}{2}$ sec. after the commencement of the motion. Camb.

Choose a suitable scale, say 1 in. to 1 sec. horizontally, and 1 in. to 20 ft. vertically. The diagram then appears as shown below. The 'explanation' asked for is as given on p. 4.

$1\frac{1}{2}$ sec. from the commencement the position of the body is represented by A. At A draw the tangent AP to the curve, cutting Y in P. Complete the right-angled triangle APQ.¹

Required velocity = AQ
 $PQ = 48 \text{ ft.} \div 1\frac{1}{2} \text{ sec.} = 32 \text{ ft./sec.}$

The second position ($2\frac{1}{2}$ sec.) is represented by B. We draw a tangent BP', but in this case the vertical line that would correspond to AQ vanishes, the slope is zero—and so the velocity is 0.

At this point you should work out an example for yourself, say no. 1 on p. 9.

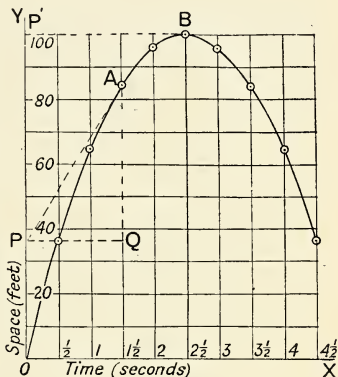


FIG. 1/4

Component velocities. Suppose a ship is steaming north very gently at 9 ft. per sec., while a man on the deck is walking north at 6 ft. per sec. At what rate is the man passing some fixed point, such as a buoy anchored in the sea?

A moment's thought shows that the answer is 15 ft. per sec. We say that the 9 ft. per sec. and the 6 ft. per sec. are the *component velocities* of the man's motion, and that 15 ft. per sec. to the north is his *resultant velocity*.

When the man has walked to the end of the deck and begins

¹ In Fig. 1/3 the triangle (ATC) was to the right of A and above it, while in Fig. 1/4 it is to the left and below. This is simply a matter of convenience, and makes no difference to the answer.

to walk at the same pace in the opposite direction, his resultant velocity will be $9 - 6 = 3$ ft. per sec. to the north, the component velocities this time being 9 ft. per sec. to the north and 6 ft. per sec. to the south.

When from two (or more) given component velocities we find the resultant velocity, we are said to *compound* the two (or more) given velocities. From the illustration just given we see that when the given velocities are in the same straight line, the process of compounding is a simple matter of addition or subtraction.

Suppose the man has become tired of walking up and down the length of the deck and begins to walk across, in an easterly direction, at right angles to the length. What is his resultant velocity now? The problem is evidently not quite so simple

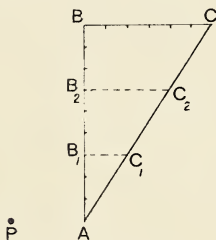


FIG. 1/5

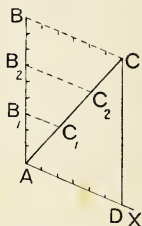


FIG. 1/6

We will suppose that when he begins his cross-deck walk he is at A, exactly opposite a fixed buoy P. If he had remained standing, in 1 sec. he would have been at B, where B is 9 ft. to the north of A. Actually, however, his walk has placed him 6 ft. to the right of B, at C. In 1 sec., then, as viewed from the fixed point P, he has moved from A to C.

Has his movement from A to C been along a straight line? Without aiming at a rigid proof, we can obtain a good idea of the answer by considering smaller intervals— $\frac{1}{3}$ sec. In this time the ship has carried him to B₁, where AB₁ = 3 ft., and his walk has placed him 2 ft. to the right of this, i.e. at C₁. Similarly, after $\frac{2}{3}$ sec. he is at C₂. If we make a drawing to scale, we find that AC₁C₂C is a straight line.

Lastly, let us consider the general case in which the two component velocities are at *any* angle—the ship moving north as before, the man travelling in a direction parallel to AX. By the time the ship would have conveyed the man 9 ft. to B (if he had been standing still), his actual position is at C where $BC = 6$ ft. and is parallel to AX. It can be shown as before that the man's movement from A to C has been along a straight line.

Our results may be summarised in the following important statement, which is known as the **Triangle of Velocities**. *If a body has two component velocities represented by the sides AB and BC of a triangle, then its resultant velocity is represented by the third side AC.*

The same truth may be expressed in another form which is sometimes convenient. Draw CD parallel to BA, cutting AX in D, thus completing the parallelogram ABCD. As before, AB represents one of the two velocities. This time, however, we will take AD instead of BC as representing the other (obviously AD is equal to BC, and is in the same direction). We may now make the following statement, known as the **Parallelogram of Velocities**. *If a body has two component velocities represented by the adjacent sides AB and AD of a parallelogram, then its resultant velocity is represented by the diagonal AC.*

Let us now consider one or two examples.

Example 2. *A man swims across a river 90 yds. wide, and when he reaches the other side he finds he has been carried 30 yds. downstream. The man's speed in still water is 2 miles per hour. At what speed is the stream flowing?*

Suppose the man starts from A. He swims at right angles to the stream, but instead of reaching the opposite point B ($AB = 90$ yds.), he finds he has been carried down to C ($BC = 30$ yds.).

Thus while the man is swimming 90 yds. the stream flows 30 yds. The speed of the stream is therefore one-third of that of the man $= \frac{1}{3}$ of 2 m.p.h. $= \frac{2}{3}$ m.p.h.

N.B. Of course the man has covered the distance AC, which is more than 90 yds., but of this, *his* contribution is only AB, while BC is that of the stream.

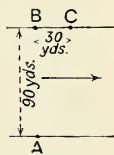


FIG. 1/7

Example 3. Continuing the last example, find the direction in which the man must swim in order to reach B, and find how long it will take.

An observer would notice that the swimmer is heading in the direction AD (Fig. 1/8), but that he is actually moving in the direction AB. Thus AB may stand for the resultant velocity of (i) the man's own velocity of 2 m.p.h. in the direction AD and (ii) the stream's velocity of $\frac{2}{3}$ m.p.h. in the direction DB.

To obtain the angle DAB, draw DB = $\frac{2}{3}$ (of any unit—say 6 cm., in which case DB would be 4 cm.) and make $\angle DBA = 90^\circ$. From centre D and with radius 2 (of the same unit) describe an arc cutting BA in A. By measurement, $\angle DAB = 19\frac{1}{2}^\circ$. Or by trigonometry, $\angle DAB = \sin^{-1} \frac{\frac{2}{3}}{2} = \sin^{-1} \frac{1}{3} = 19\frac{1}{2}^\circ$.

To find the man's time, we can now let AB (in the figure we have just drawn) stand for the width of the stream, 90 yds. By measurement or calculation we find that AD = 95.4 yds. In still water the man would have had to swim this 95.4 yds. at 2 m.p.h., and by arithmetic we find that the required time would be **1 min. 38 sec.**

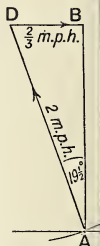


FIG. 1/8

Relative velocity. Closely connected with the subject of component velocities, etc., is that of relative velocity. We may say that **the relative velocity of body A with respect to another body B** is the velocity with which A appears to be moving when viewed from B.

If a train travelling north at 60 m.p.h. is passing a telegraph pole, the pole appears to a passenger to be travelling south at 60 m.p.h. This is the relative velocity of the telegraph pole with respect to the passenger.

If at a junction a train A moving west at 12 m.p.h. is overtaking another train B at 8 m.p.h., the velocity of A relative to B is 4 m.p.h. to the west. If this is not quite clear, draw a diagram for yourself and you will soon see it.

As another example, take the passenger walking across the deck (p. 6, Fig. 1/5). His velocity relative to another (stationary) passenger on deck is 6 ft. per sec. to the east, while his velocity relative to a seagull perched on P would be about 10.8 ft. per sec. in a direction some 34° east of N.

An aeroplane pilot is provided with an air-speed indicator which gives him his speed *relative to the air*. If this becomes too low he is in danger of stalling and then spinning downwards. Accidents have sometimes happened because a pilot flying

with the wind has adjusted his speed with reference to the ground. For instance, if he is flying with a 50 m.p.h. wind and his speed relative to the ground is 80 m.p.h., his speed relative to the air—the point that matters—is only 30 m.p.h. Here is an example involving relative velocity.

Example 4. A ship is sailing north-west with a velocity of 10 p.h. A passenger on board notices that the wind seems to blow from the north with a velocity of 15 m.p.h. Find the true direction and velocity of the wind.

Consider some smoke coming from the funnel. The motion of the ship would give this an apparent velocity of 10 m.p.h. from the north-west. The wind observed by the passenger is the resultant of :

- (i) the 10 m.p.h. movement from the north-west just noted and
- (ii) the actual velocity of the wind.

We therefore draw AB to represent this resultant (15 m.p.h.) and AC to represent (i). The actual velocity of the wind is evidently represented by CB. By measurement we find that this is **10.6 m.p.h.**, and that it is blowing from direction **42° degrees east of north.**

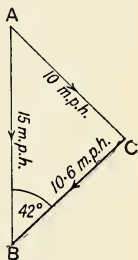


FIG. 1/9

Questions

A

1. A man walks at a uniform pace of 4 m.p.h., but rests for quarter of an hour at the end of the second hour. Draw a space-time diagram to represent his motion.
2. The following velocities are given in miles per hour : 60, 30, 00, 120. Express them in ft. per sec.
3. How many miles per hour correspond to (i) 88, (ii) 22, (iii) 132 ft. per sec.?
4. A ship is steaming along a canal at 6 m.p.h., and a man on the ship is walking in the same direction at $3\frac{1}{2}$ m.p.h. What will be the man's velocity relative to a stationary observer on the shore? What will this velocity be when the man turns round and walks in the opposite direction?

B

1. A body is projected vertically upwards, and its distance from the ground after $\frac{1}{2}$, 1, $1\frac{1}{2}$, 2, $2\frac{1}{2}$, 3, $3\frac{1}{2}$ and 4 sec. are found to be 28, 48, 60, 64, 60, 48, 28 and 0 ft., respectively. Represent the motion by a space-time diagram, and from the latter find the velocity of the body after it has been moving (i) 2 sec. (ii) $3\frac{1}{2}$ sec.

2. An aeroplane is to travel from A to another place B 200 miles due north of it. Assuming the wind to remain constant at 35 miles per hour directly from the east, and the aeroplane to have a speed of 125 miles per hour in still air, determine (i) the direction in which the aeroplane has to be steered, (ii) the time taken to travel the 200 miles. *C.W.B.*

3. What is meant by the *parallelogram of velocities*?

An aeroplane which has a speed of 250 miles per hour, in still air, travels from a point A to a point B which is 400 miles due east of A. If the wind is blowing from the north at 50 miles per hour, determine (a) the direction in which the aeroplane has to be steered, and (b) the time taken to travel from A to B. *C.W.B.*

4. Four ships in convoy are at the corners of a square with side one mile long. Ship B is due north of A; C is due east of B, and D is due south of C. The four ships are each sailing north at 10 miles per hour. A motor-boat which travels at 26 miles per hour starts at A, delivers a message to B, then to C, then to D, and finally returns to A. The distances between the ships are traversed by the shortest possible routes, and no allowance need be made for time taken in turning.

How far does the convoy advance whilst the motor-boat is going round?

Reprinted by permission from The Sunday Times, in which the question appeared some years ago as a holiday competition.

5. Define (a) velocity, (b) acceleration.

A mass on a smooth table moves through the following distances in the stated times. Draw a space-time curve with the times on the X-axis, and from it deduce the velocity at time 5.0 sec.

Total time (sec.).	0	1.0	2.0	3.0	4.0	5.0	6.0	7.0
Total distance (cm.)	0	10.0	41.0	91.0	162	253	364	495

Camb.

CHAPTER 2

ACCELERATION. FALLING BODIES

It may be that you have had the alarming experience of descending a steep hill on a bicycle when something has gone wrong with the brakes. If so, you will have gained a very good idea of the meaning of the term *acceleration*, though you would certainly not be in a state of mind to attempt an exact definition. We will try to arrive at one now, but perhaps two motor-cars will serve our turn better than a runaway bicycle. Two cars A and B both start from rest. In 10 sec. A has acquired a velocity of 30 m.p.h., but it takes B 20 sec. to acquire this same velocity. A, then, has increased its velocity by 30 m.p.h. in 10 sec., so that its average rate of increase is 3 m.p.h. per sec. B has increased its velocity by 30 m.p.h. in 20 sec., its average rate of increase being $1\frac{1}{2}$ m.p.h. per sec. Because A is increasing its velocity at a more rapid rate than we say that its acceleration is greater. In fact *the acceleration of a body is the rate at which its velocity is increasing*.

It follows from our definition that if the velocity is decreasing, the acceleration is negative. A negative acceleration is often known as a *retardation*.

Acceleration is said to be *uniform* when the velocity changes equally in equal intervals of time, however short these intervals may be. In the present chapter it is to be understood that acceleration is uniform, unless something is said or implied to the contrary.

Example 1. A train increases its velocity from 10 m.p.h. to 30 m.p.h. in 5 mins. Find its average acceleration (i) in m.p.h. per min., (ii) in ft. per sec. per sec.

- (i) The velocity has increased by 20 m.p.h. in 5 mins. This is at the rate of **4 m.p.h. per min.**
 (ii) $4 \text{ m.p.h. per min.} = \frac{88}{15} \text{ ft. per sec. per min.}$
 (for $60 \text{ m.p.h.} = 88 \text{ f.p.s.}$, and $4 \text{ m.p.h.} = \frac{1}{15}\text{th}$ of this).

If $\frac{88}{15} \text{ ft. per sec.}$ is the increase of velocity per *minute*, the increase per *second* will be $\frac{1}{60}\text{th}$ of this.

$\therefore \text{acceleration} = \frac{1}{60} \times \frac{88}{15} = 0.098 \text{ ft. per sec. per sec.}$

Notice how the answers are expressed, (i) miles per hour per minute; (ii) feet per second per second. People are sometimes puzzled to find that a unit of time is mentioned twice. It is mentioned first, of course, to indicate how much velocity has been added on (e.g. '4 miles per hour'), and afterwards to give the rate at which that increase has been taking place ('per minute'). For shortness, 0.098 ft. per sec. per sec. would often be written 0.098 ft./sec.²

Example 2. A train travelling at 60 m.p.h. is brought to rest by the action of its brakes in 2 minutes. Find the average acceleration in ft./sec.²

Here the acceleration is obviously *negative*.

$$60 \text{ m.p.h.} = 88 \text{ f.p.s.}$$

$$\therefore \text{loss of velocity} = 88 \text{ ft. per sec. in 2 mins.}$$

$$= \frac{88}{120} \text{ ft. per sec. per sec.}$$

$$= 0.73 \quad ,, \quad ,,$$

$$\text{i.e. acceleration} = - 0.73 \text{ ft./sec.}^2$$

(We could have said *retardation* = 0.73 ft./sec.², but this is less usual.)

Velocity-time curves. On p. 3 we made a displacement-time (or space-time) curve. Let us now make a few velocity-time curves and see what can be learnt from them.

We will begin with the case of a body moving with a constant velocity (i.e. acceleration = 0) of 5 ft. per sec., for 6 sec.

The 'curve' is a straight line parallel to the time-axis, i.e. the slope = 0. This is because there is no change in the velocity, i.e. *acceleration* = 0. We shall see presently that the slope is always a measure of the acceleration.

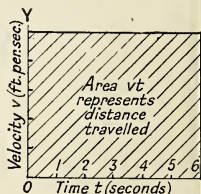


FIG. 2/1

The distance travelled by a body moving for 6 sec. at 5 ft. per sec. would evidently be $6 \times 5 = 30$ ft. On our graph this distance is represented by the shaded area under the curve.

We will next consider the velocity-time curve for a body

moving with a constant acceleration (other than zero). Here we have the necessary data for a body thrown upwards with a velocity of 100 ft. per sec. (we shall see later that the acceleration in such a case is constant and $= -32 \text{ ft./sec.}^2$).

Time (sec.)	0	$\frac{1}{2}$	1	$1\frac{1}{2}$	2	$2\frac{1}{2}$	3	$3\frac{1}{2}$	4
upward velocity (ft./sec.)	100	84	68	52	36	20	4	-12	-28

Constructing the curve, we have the figure given below:—

(i) The first obvious point is that our 'curve' is again a straight line. This is a mathematical consequence of the fact that for equal changes in the time (0, 1, $1\frac{1}{2}$. . .) we have equal changes in the velocity (100, 84, 68, . . .), i.e. of the fact that the acceleration is uniform.

(ii) In our space-time graph (p. 3) we saw that from the slope we obtained the quotient $\text{change of position} \div \text{time}$, i.e. *velocity*. Similarly, from the slope of our velocity-time graph we obtain the quotient $\text{change of velocity} \div \text{time}$, i.e. *acceleration*.¹

Thus, taking the slope as $OB \div OA$, we get acceleration

Thus in Example 1 (p. 11) we obtained the acceleration (4 m.p.h. per min.) by dividing the change of velocity (20 m.p.h.) by the time (5 mins.).

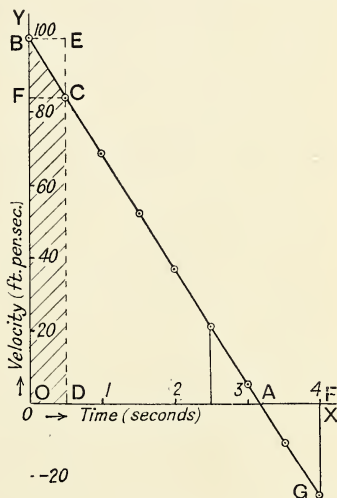


FIG. 2/2

$= -100 \text{ ft./sec.} \div 3\frac{1}{8} \text{ sec.} = -32 \text{ ft./sec.}^2$ (We write *minus* 100 because the velocity is decreasing.)

(iii) Consider a small portion of the curve on the line between OB and the point C (where velocity = 84 ft./sec. and time = $\frac{1}{2}$ sec.). From C drop a perpendicular CD to the time-axis. Produce DC to E, and draw BE and parallel to OD.

At the beginning of the half-second the velocity was 100 ft./sec. If it had remained at this value the distance travelled would have been $100 \times \frac{1}{2} = 50$ ft. This would have been represented by the area OB \times OD, giving the rectangle OBE.

On the other hand, if the velocity throughout had been the same as at the *end* of the interval (84 ft. per sec., represented by DC) the distance would have been $84 \times \frac{1}{2} = 42$ ft. This would have been represented by the area OFCD.

For practically the whole interval of half a second, however, the velocity has been less than 100 but greater than 84. Therefore the distance travelled is less than 50 (the rectangle OBE) but greater than 42 (rectangle OFCD). It is actually represented by the intermediate area, OBCD¹, which in Fig. 1 is indicated by shading.

This is the distance travelled in the first half-second. Continuing in the same way, we see that the distance travelled in $3\frac{1}{8}$ sec. (which brings us to the point A) is represented by the space OAB.

What about the area AGF (between $3\frac{1}{8}$ sec. and 4 sec.)? This lies *below* OX, indicating that the velocity throughout the interval is *negative*.

This corresponds to the physical fact that the body thrown upwards has reached its highest point (at A, velocity = 0) and is now descending. If its upward velocity is reckoned positive, clearly its downward velocity is negative. Hence the area AGF represents a negative distance—the moving body is ‘turning back on its tracks.’

The actual distance travelled in the first 4 sec. is represented then, by the area OAB *minus* the area AGF.

¹ This is not a rigid proof. By taking DC infinitely close to OD, however, it can be proved that the space described is really represented by the area OBCD, even though the complete ‘curve’ is not a straight line.

As a last velocity-time curve let us take a case in which acceleration is variable. We have abundant illustrations the movement of a motor-car. Thus from rest to 30 m.p.h. we usually have rapid acceleration, while from 50 m.p.h. to 60 m.p.h. acceleration would be much slower. If a car pulls up sharply, acceleration is very rapid, but negative. If it maintains a steady velocity of 40 m.p.h. some distance, acceleration is zero, and so on. Let us take the following case.

Time (sec.)	0	1	2	3	4	5	6	7	8
Velocity (ft./sec.)	0	10	17	23	28	32	35	37	38

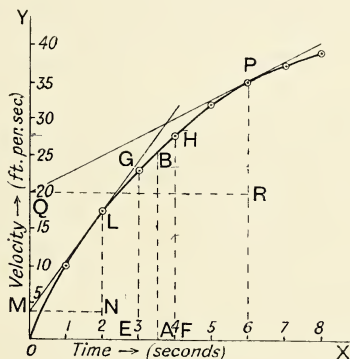


FIG. 2/3

It will give point to our study of this curve if we try to answer the following questions:—

- (i) What is the velocity at $3\frac{1}{2}$ sec. from rest?
- (ii) How far did the body travel in the fourth second?
- (iii) How far did it travel in the first 4 sec.?
- (iv) What is its acceleration (a) 2 sec., (b) 6 sec. after starting?

- (i) Mark the point $3\frac{1}{2}$ sec. (A), draw the perpendicular AB and measure it. It indicates a velocity $25\frac{1}{2}$ ft. per sec.
- (ii) Mark the end of the third second (E) and of the fourth (F). From E and F draw perpendiculars EG and FH to the time-axis, cutting the curve at G and H. The area GEFH gives the distance travelled in the fourth second. This area may be obtained either by calculation—it is practically equal to $EF \times \frac{1}{2}(GE + HF)$ —or by counting squares, and indicates a distance of $25\frac{1}{2}$ ft.
- (iii) The distance travelled in the first 4 sec. is represented by the area OFH and (by counting squares or otherwise) is found to be 64 ft.
- (iv) The acceleration 2 sec. after starting is found by drawing the tangent LM and measuring the slope $LN \div MN = 12\frac{1}{2} \div 2 = 6\frac{1}{4}$ ft./sec.²

Similarly the acceleration 6 sec. after starting = PR
 $QR = 15 \div 6 = 2\frac{1}{2}$ ft./sec.²

Falling bodies. In the opening paragraph of this book we saw that the subject of falling bodies forms a landmark in the history of science. Galileo proved that in this matter the ideas of Aristotle were wrong, and if he was wrong in one point he might be wrong in others. The last word in science had evidently not been said, and men felt encouraged to make experiments for themselves with a view to understanding the hidden laws of nature.

But though Galileo had found what the law of falling bodies was *not*, he was far from having found what it *was*. Clearly the farther the body fell, the greater was the velocity, and at first he asked himself whether the velocity might be proportional to the distance fallen—whether, for instance, a body after falling 100 ft. would have twice the velocity that it had after falling only 50. By a process of reasoning without actual experiment, he satisfied himself that this was not the case.

If the velocity acquired is not proportional to the distance, is it proportional to the time? In other words, is the acceleration uniform? Galileo was able to prove mathematically

It if this is the case, *then* the distance fallen would be proportional to the square of the time; that, for instance, a body would fall nine times as far in 3 sec. as it would in 1 sec. This was the point he had to settle—by experiment.

The difficulties were very great. A falling body soon begins to move so rapidly that accurate timing is nearly impossible even with a modern stop-watch—and the only watches available in Galileo's day were very clumsy affairs.

He got over the difficulty of rapid movement in this way.

'I made the body run down an inclined plane instead of falling vertically,' he argued, 'the whole movement would be slowed down, but the same law would be at work—the distance travelled would still be proportional to the square of the time.'

He proposed, so to speak, to *dilute* the force of gravity.

But without a good watch how was he to measure time?

He filled a large vessel with water and made a pin-hole in the bottom, so that the water could drip out. He collected the water in a weighed dish, and then the weight of the water was a measure of the time.¹

So he took a board about 23 ft.² long, and in it made a groove about an inch wide.³ The groove was straight and smooth, and was covered with very smooth parchment.

He tilted the board slightly, and found the time required for a polished brass ball to roll down, for instance, (a) the full length of the board, (b) a quarter of that length. He found

that the first time was just twice the second, i.e. the distance travelled was proportional to the square of the time. To

quote his own words, 'We repeated this experiment more than once, in order to measure the time with an accuracy such that the deviation between two observations never exceeded one-tenth of a pulse-beat . . . in such experiments,

repeated a full hundred times, we always found that the spaces traversed were to each other as the squares of the times, and this was true for all inclinations of the plane. . . .

Provided the level of the water in the large vessel was kept constant. Galileo does not say whether he attended to this point.

'12 cubits.' Torricelli gives the height of his barometer, which would be about 30 in., as 'a cubit and a quarter and an inch over.' This is an interesting little exercise in arithmetic to calculate, from this, the equivalent of 12 cubits.

'A little more than one finger in breadth.'

The water thus collected was weighed, after each descent on a very accurate balance; the differences and ratios of the weights gave us the differences and ratios of the times, and this with such accuracy that although the operation was repeated many, many times, there was no appreciable discrepancy in the results.'¹

Modern forms of experiment. The obvious modern form of Galileo's experiment is to use a stop-watch or stop-clock instead of his dripping-water method. A grooved board 8 or 10 ft. long can be prepared without great difficulty, and marked off into lengths of 1 ft. with pencil or chalk. With a little practice, the times required to traverse 1, 2, 3 . . . can be found with very fair accuracy, the time in each case being taken as the mean of several values. The results could evidently be expressed as a space-time curve (p. 10), but it will be more to our present purpose to plot the 'space' against the *square* of the time.

Thus, suppose we have obtained results corresponding to the first two rows in the following table. We find t^2

Length (ft.).	1	2	3	4	5	6	7	8	9	10
Time (t) (in sec.)	3	$4\frac{1}{5}$	$5\frac{1}{5}$	6	$6\frac{3}{5}$	$7\frac{2}{5}$	8	$8\frac{3}{5}$	9	$9\frac{4}{5}$
t^2	9.0	17.6	27.0	36.0	43.6	54.8	64.0	70.6	81.0	88.0

arithmetic, and then obtain the graph shown in Fig. 2/4. The fact that the graph is a straight line proves that the distance travelled is proportional to the *square* of the time, and this in turn requires that the acceleration shall have been uniform. This uniform acceleration to which a freely falling body is subject is known as g .

How Aristotle was misled. We have all noticed that a penny

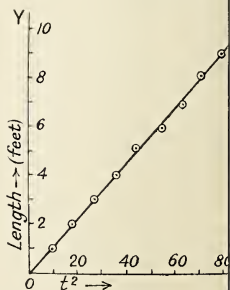


FIG. 2/4

¹ Quoted from Magie, *Source Book in Physics*.

is to the ground a great deal faster than a piece of paper, and that very light particles such as dust fall very slowly indeed. We can therefore hardly blame Aristotle for concluding that heavy bodies fall more quickly than light ones, for he was certainly taking far too much for granted when he assumed that the rate of fall is actually *proportional* to weight.

The point he overlooked was the part played by the *air*. If it were not for this, a piece of paper would fall to the ground just as quickly as a penny.

To prove it, cut out a circle of paper a shade smaller than a penny. Lay it on the coin, hold the latter horizontally between thumb and finger, and then let it drop. Paper and penny both reach the floor at the same instant.

If the penny had been uppermost, we might have argued that it had simply pushed the paper in front of it; but the penny was underneath.

What has happened is that the penny has pushed the air out of the way, and then the paper, relieved of the resistance of the air, has fallen just as fast as the penny.

Boyle placed a leaden bullet and a piece of paper in the receiver of his air-pump—the receiver in this case being a glass tube—and pumped out the air. He then quickly turned the tube upside down, and noted that bullet and paper reached the bottom at the same time. Newton did a similar experiment with a guinea and a feather. Experiments of the same sort can easily be done in school if an air-pump is available.

The reason why heavy bodies and light bodies are all subject to the same acceleration will be better understood when we come to the next chapter, but we can obtain a rough idea now. Suppose a penny weighs 100 times as much as a circle of paper. In that case the force pulling it downwards is 100 times the force acting on the piece of paper, and we feel that it ought to move faster.

But is it not possible that the penny contains 100 times as much *stuff* or *matter* as the paper? In that case, the larger force acting on the larger amount of matter, might produce exactly the same acceleration as the smaller force acting on the smaller amount.

Determination of g . The value of g can be found very accurately by means of the pendulum, with which you have probably made experiments. It consists of a heavy bob tied to the end of a piece of thread, and you will have found that the time of swing does *not* depend on the weight of the bob, nor on the amplitude of the swing (i.e. on the distance from the central position of the bob to the extreme left or right). It *does* depend, however, on the length, the time being proportional to the square root of the length ($t \propto \sqrt{l}$). It can be proved in fact that

$$t = 2\pi \sqrt{\frac{l}{g}}$$

where t is the time of swing in seconds, $\pi = 3.14$, l is the length of the pendulum in centimetres, and g is the acceleration due to gravity, in cm. per sec. per sec.

According to this equation, an increase in the value of g would cause t to be reduced. This seems to be reasonable. For suppose the bob to be at A in its extreme left position. It begins to move along the path AV, and the speed with which it moves along this downward path will depend on the acceleration due to gravity. The greater this is (i.e. the greater the value of g), the more quickly will the bob move downwards (i.e. the smaller the value of t).

So to find the value of g , we make a pendulum and measure l (SV in Fig. 2/5) in centimetres. We then take the time of (say) 20 swings in seconds, and dividing by 20 we obtain the value of t . Knowing t , π and l , we can obviously calculate g .

Actually, the value of g at Greenwich is 981.18 cm. per sec. per sec. or 32.19 ft. per sec. per sec. In all ordinary calculations the values are taken as 981 and 32, respectively, and these are the only numbers in this paragraph that need be remembered.

It is interesting to notice that at the equator $g = 978$ while at the North Pole it is 983.21. This is because, owing to a slight flattening at the poles, the distance from the earth's centre is a little less at those points, and so the force of grav-



FIG. 2/5

a little greater. Similarly, while the value is 981.18 at Greenwich, it is 981.37 at Manchester, 200 miles farther north. For the same reason (i.e. increased distance from the earth's centre), the value of g might be expected to fall very slightly with increased height above sea-level, and this is found to be so.

Uniformly accelerated motion—Equations. Suppose that at a certain instant a body has a velocity u (represented by OA in the velocity-time graph of Fig. 2/6) and that it has a uniform acceleration a .

After 1 sec. its velocity will be $u + a$.

„ 2 sec. „ „ $u + 2a$.

„ 3 sec. „ „ $u + 3a$.

„

„

„ t sec. „ „ $u + ta$.

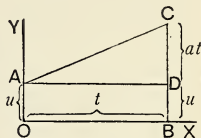


FIG. 2/6

this velocity is denoted by v , we may write $v = u + at$. (1) It is represented in our graph by the line BC .

Draw AD parallel to the time axis cutting BC in D .

Then $BD = OA = u$.

$\therefore DC = BC - BD = (u + at) - u = at$.

The velocity-time graph is AC , and as the acceleration is uniform, AC is a straight line.

The distance s described in time t is represented by the area $OACB$ (p. 14).

Now area $OACB = \text{rect. } OADB + \triangle ADC$

$$= (OA \times OB) + \frac{1}{2} AD \times DC$$

$$= (u \times t) + \frac{1}{2} t \times at^*$$

$$\text{i.e. } s = ut + \frac{1}{2} at^2 \quad (2)$$

s may be expressed in another way, for $DC = BC - BD = v - u$. Giving this value to DC , the line marked with an asterisk becomes $(u \times t) + \frac{1}{2} t \times (v - u)$

$$\text{i.e. } s = ut + \frac{1}{2} vt - \frac{1}{2} ut$$

$$= \frac{1}{2} ut + \frac{1}{2} vt$$

$$\text{or } s = \frac{1}{2} (u + v)t \quad (3)$$

Lastly, from (1) we have $v - u = at$

$$\text{and } \quad (3) \quad v + u = \frac{2s}{t}$$

Multiplying together, we have $v^2 - u^2 = 2as \quad (4)$

Equations 1, 2, 3 and 4 can be used to solve any problem on uniformly accelerated motion, and they should be carefully committed to memory.

Motion from rest. In the case just discussed, the body was already travelling with velocity u when we began to take notice of its motion. Very often, however, it happens that we are considering the uniformly accelerated motion of a body starting *from rest*. In that case $u = 0$ and our equations (1), (2), (3) and (4) become respectively

$$v = at \quad . \quad . \quad . \quad . \quad . \quad . \quad (1a)$$

$$s = \frac{1}{2} at^2 \quad . \quad . \quad . \quad . \quad . \quad . \quad (2a)$$

$$s = \frac{1}{2} \check{a} t^2 \quad . \quad . \quad . \quad . \quad . \quad . \quad (3a)$$

$$\text{and } v^2 = 2 \check{a} s \quad . \quad . \quad . \quad . \quad . \quad . \quad (4a)$$

We will now work through a few typical examples of uniformly accelerated motion.

Example 1. *A body falls from rest for $1\frac{1}{2}$ sec. What velocity has it acquired, and how far will it have fallen?*

We have $u = 0$, $a = 32$, $t = 1\frac{1}{2}$, $v = ?$ $s = ?$

Using the equation $v = u + at$

$$\begin{aligned} \text{we have } v &= 0 + 32 \times 1\frac{1}{2} \\ &= 48 \text{ ft. per sec.} \end{aligned}$$

Also from $s = \frac{1}{2} (u + v)t$

$$\begin{aligned} \text{we have } s &= \frac{1}{2} (0 + 48) \times 1\frac{1}{2} \\ &= 36 \text{ ft.} \end{aligned}$$

Sometimes we can manage quite well without a formula. In this case for instance,

Acceleration = 32 ft. per sec. per sec.

\therefore in $1\frac{1}{2}$ sec., velocity acquired = $1\frac{1}{2} \times 32 = 48$ ft. per sec.

As the velocity of the body is 0 to begin with, and increases uniformly to 48 ft. per sec., the *average* velocity is $\frac{1}{2}(0 + 48) = 24$ ft. per sec.

Distance fallen in $1\frac{1}{2}$ sec. at an average velocity of 24 ft. per sec. = $1\frac{1}{2} \times 24 = 36$ ft.

Example 2. *A body is thrown upwards with a velocity of 80 ft. per sec. How far will it rise? Find its velocity (a) when half the time has elapsed, (b) when it has attained half its full height.*

The first two answers are easily obtained without the use of formulæ, thus:—

The body *loses* a velocity of 32 ft. per sec. every sec.

∴ to lose the whole of its velocity, no. of sec. required

$$= 80 \div 32 = 2\frac{1}{2}.$$

Its average velocity will be $\frac{1}{2}(80 + 0) = 40$ ft./sec.

∴ height attained = (average vel.) \times (time)

$$= 40 \times 2\frac{1}{2}$$

$$= 100 \text{ ft.}$$

When half the time has elapsed, it will have lost half its velocity, which will therefore be reduced to **40 ft. per sec.**

For the last part, it is easier to use a formula.

We have $u = 80$. $s = 50$. $a = -32$. $v = ?$

The formula connecting these quantities is $v^2 = u^2 + 2as$.

$$\therefore v^2 = 80^2 + 2(-32) \times 50$$

$$= 6400 - 3200$$

$$= 3200$$

$$\therefore v = \sqrt{3200} = 56.6 \text{ ft. per sec.}$$

(Although acceleration here is negative, the formula is unaltered. It still reads $v^2 = u^2 + 2as$, but we substitute -32 for a .)

Example 3. A stone was thrown vertically upwards from the top of a cliff. After 8 sec. it passed the top of the cliff as it descended, and 3 sec. later fell into the sea below. Find the height of the cliff.

The stone has evidently taken 4 sec. to reach its highest point, its initial velocity is reduced to zero in 4 sec.

But loss of velocity per sec. = 32 ft. per sec.

∴ initial velocity (upwards) must have been $4 \times 32 = 128$ ft. per sec.

This could also be easily obtained from the formula $v = u + at$, putting $v = 0$, $a = -32$, $t = 4$, $u = ?$).

As it passes the cliff, the stone will have a *downward* velocity 128 ft. per sec. and 3 sec. later this will have increased to $128 + (32 \times 3) = 224$ ft. per sec.

∴ average velocity on this journey from cliff to sea = $\frac{1}{2}(128 + 224) = 176$ ft. per sec.

The stone travels with this average velocity for 3 sec.

∴ height of cliff = $176 \times 3 = 528$ ft.

(Or we might say, $u = 128$, $a = 32$, $t = 3$, $s = ?$ and use the formula $s = ut + \frac{1}{2}at^2$.)

Object thrown horizontally. It is an interesting exercise to trace the path of a stone thrown *horizontally* from the edge of a high cliff with a given velocity. Suppose the latter

is 40 ft. per sec. and we wish to find the position of the body after $\frac{1}{2}$, 1, $1\frac{1}{2}$, 2 . . . sec., air resistance being neglected.

We will work out the '2 seconds' case, because as numbers are fairly large, they will show more clearly in diagram.

A is the point from which the body is thrown. Its initial downward velocity is 0, and from the formula $s = ut + \frac{1}{2}at^2$ we easily calculate that if there had been no outward motion it would have reached B_1 , where $AB_1 = 64$ ft.

Its horizontal velocity is 40 ft. per sec. (unchanging, if we neglect air resistance), so after 2 sec., if there had been no downward motion, it would have reached B_2 , where $AB_2 = 80$ ft. Clearly its actual position will be at B, 64 ft. below B_2 .

Work out the other positions for yourself. You should obtain a curve such as is shown in Fig. 2/7.

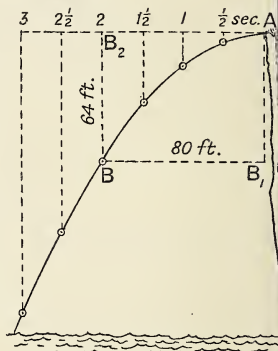


FIG. 2/7

Questions

A

1. A train increases its velocity from 10 m.p.h. to 60 m.p.h. in 10 mins. Find its average acceleration, (i) in m.p.h. per minute (ii) in ft. per sec. per sec.

* 2. A stone is falling with an acceleration of 32 ft./sec.². Express this acceleration in ft. per sec. per minute.

3. A body projected vertically upwards with a velocity of 80 per sec. has the following velocities at the times mentioned:—

Time (sec.).	0	$\frac{1}{2}$	1	$1\frac{1}{2}$	2	$2\frac{1}{2}$	3	$3\frac{1}{2}$	4	$4\frac{1}{2}$	5
Upward velocity (ft./sec.)	80	64	48	32	16	0	-16	-32	-48	-64	-80

Construct the velocity-time curve, and from it find (i) acceleration, (ii) the distance travelled in $2\frac{1}{2}$ sec.

B

1. Explain the terms *velocity*, *uniform velocity*, *average velocity*. A motor-car moving with a velocity of 24 ft./sec. is uniformly accelerated at 2 ft./sec.² for 20 sec., and is then uniformly retarded as to come to rest after a further 16 sec. Plot the velocity-time graph of the motion, and use the graph to determine (a) the total distance travelled, (b) the average velocity of the car. O.C.

2. A constant braking force is applied to a train, so reducing uniformly its velocity of 44 ft. per sec. by 8 ft. per sec. every 10 sec. Compile a table showing at the end of successive 10 sec. up to 50 sec. (a) the velocity, (b) the total distance travelled by the train from the application of the brakes.

On the same axes plot a velocity-time graph and a distance-time graph, and on them indicate (i) the time taken to come to rest if the brakes are kept on, (ii) the distance travelled during that time. Also determine these values by calculation. Lond.

3. Explain how you would use the velocity-time graph for a moving body to deduce (a) the acceleration at any instant, (b) the space described during any interval of the period for which the graph is drawn.

A stone is thrown vertically upwards with a velocity of 80 ft./sec. Plot a velocity-time graph for the first 5 sec. of the motion, and find from the graph the greatest height reached. ($g = 32$ ft./sec./sec.) Oxf.

4. Define uniform acceleration and describe an experiment to find g , the acceleration due to gravity.

A body is projected vertically upwards from the ground and reaches a height of 625 ft. Find the velocity of projection and the time to rise to the highest point. Lond.

5. An aeroplane flying horizontally at 240 miles per hour at a height of 5000 ft. releases a bomb. Neglecting air resistance, calculate (a) the time which elapses before the bomb hits the ground, and (b) the horizontal distance travelled by the bomb before the impact. O.C.

6. A shell is fired vertically with a velocity of 400 ft. per sec. How high will it rise, and after what time will it reach this maximum height? O.C.

7. A cricket ball is thrown vertically upwards with a velocity of 96 ft. per sec. Calculate how far it will rise and how long will it take before it again hits the ground.

8. From a balloon ascending with a vertical velocity of 16 ft./sec., a stone is let fall which reaches the ground in 10 sec. Neglecting friction, find the height of the balloon when the stone is released.

CHAPTER 3

NEWTON'S SECOND LAW OF MOTION

WE would all agree that if a brick (or any other object) lying on the ground, it will remain exactly where it is unless some force is applied to it. Perhaps we are not so sure that if a body is *moving* in a straight line—a cricket ball rolling over the ground, for instance—it will continue moving forever, with the same velocity, unless some force is applied to it. Our cricket ball soon comes to rest because of the force of *friction*. If we could make a 'cricket ball' of smooth ice and bowl it along the smooth surface of a frozen lake, it would continue to move for a much longer time, because we have greatly reduced the force of friction. But we cannot get rid of this force altogether—and there is always the frictional resistance of the air. However, thinking along these lines, it is not hard to realise that if our moving cricket ball could be set *completely* free from resisting forces it would continue to move for ever with unchanging velocity. The ideas we have thus briefly discussed are summed up in Newton's **First Law of Motion**, which states that *A body in a state of rest or of uniform motion in a straight line will continue that state of rest or motion unless it is acted upon by some external force.*

In short—no force, then no change of motion.

If change of motion is produced only by force, we ought somehow to be able to measure force by the change of motion it produces. Let us follow up this idea and see where it leads us.

Two boys of equal size are in two quite similar rowing boats. They start together, working their hardest, and after 30 sec. A has attained a speed of 3 m.p.h. B takes 50 sec. to acquire the same speed. Quite rightly, we conclude that A has exerted more force than B. Force, then, is measured not simply by the velocity it imparts to a body, but by the *rate* at which the velocity is imparted—in short, by the acceleration.

Let us start A and B off once more, but this time we will provide A with a good heavy cargo, so that he has to propel much greater mass than B. Once again both row their hardest, but this time A takes 50 sec. to acquire a speed of m.p.h.—the same time as B.

In this case also we conclude that A has exerted more force than B, because though he has produced only the same acceleration, his force has been applied to a larger mass.

If, then, we are to estimate force by the motion it produces, we must take into account: (i) the *mass* of the body to which the force is applied, and (ii) the *rate* at which it causes the velocity of the body to increase.

Mass and velocity are very conveniently combined in the term *momentum*, which means the product of the two. Thus it seems that a force might be measured by applying it to a body and noting the rate at which the *momentum* of that body is increased. To go back to our illustration, A's rate of increase of momentum was greater in both cases. In the first he was dealing with the same mass as B, but caused its velocity to increase at a greater rate. In the second he was dealing with a greater mass, and caused its velocity to increase at the same rate.

Let us now try to obtain more exact information on the two points mentioned, to obtain, in fact, the answers to these two questions:—

- (i) What is the exact relation between the force applied to a body of given mass and the acceleration it produces?
- (ii) Suppose the same force is applied to successive bodies of different mass, what will be the relation between the various accelerations produced and the respective masses?

The apparatus most commonly used in dealing with these questions is known as *Fletcher's Trolley*, so we will begin by describing it.

The trolley itself consists of a block of wood about 2 ft. long mounted on wheels, which turn with very little friction. At right angles to its length it is drilled with holes which carry several steel cylinders S_1, S_2, \dots . These serve to weight the block, and also enable us to vary its mass when required.

The trolley runs over a base-board, the surface of which is very smooth and hard—in fact, it is usually faced with steel with grooves for the trolley wheels. A cord is attached to the trolley and passes over a pulley, afterwards being attached to a light cardboard box in which weights may be placed, causing the trolley to move.

Near the back is a fixed iron pillar to which is clamped a straight, steel spring which can be made to vibrate in a horizontal plane. By altering the setting, the vibrating portion can be made longer or shorter, and thus the period of vibration can be adjusted to $\frac{1}{5}$ th sec., $\frac{1}{8}$ th sec., etc. Near the front of the spring is a small brush.

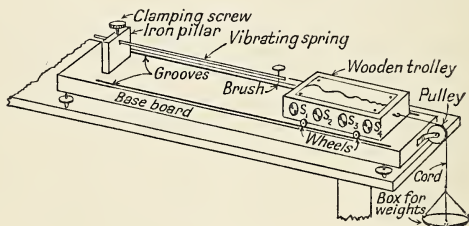


FIG. 3/1

the free end is a small brush, which is dipped in ink from time to time. A strip of paper is pinned on the top of the trolley and as the spring vibrates while the trolley is moving forward the brush traces a wavy line.

We begin our experiment by so arranging the conditions that when the trolley is made to move forward with a slight tap, it shall continue to move with uniform velocity. This is done by putting a small weight (2–5 gm.) in the box (just short of what is needed to make the trolley move as required), completing the adjustment by a turn of the rear screw of the base-board. When the adjustment is correct the wavy line traced on the paper will be of equal length (Fig. 3/2a).

We may now consider that our block is free from the action of outside forces. Of course we cannot quite eliminate friction, but the frictional forces are neutralised by the slight screw motion and by the weight we have put in the box.

We can now attempt to answer question (i)—to find

exact relation between the *force* applied to a body of given mass and the *acceleration* it produces.

We put a certain weight, say 10 gm. in the box (additional to the 2-5 gm., already there), start the trolley, and obtain a curve such as Fig. 3/2*b*.

This 10 gm. was, of course, the 'applied force.'¹

We now repeat the process using 20 gm. instead of 10 gm., and obtain Fig. 3/2*c*.

In Fig. 3/2*b* draw the centre line ABCD, and mark a point where this cuts one of the first waves. B is a corresponding point say four complete waves to the right, while C and D are respectively eight and twelve waves to the right.

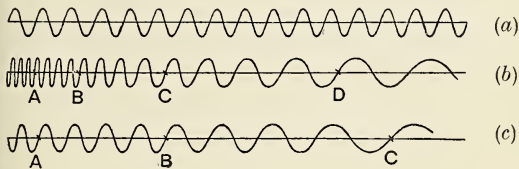


FIG. 3/2

Now the spring makes its vibrations *in equal times* (like a pendulum). Thus the trolley has moved through the distances AB, BC, and CD in equal intervals of time (namely, the time occupied by four vibrations of the spring). Measure AB, BC, and CD. In an actual experiment these were found to be 1.59 in., 2.98 in., and 4.37 in. We notice that $2.98 - 1.59 = 1.39$, and that $4.37 - 2.98 = 1.39$. Thus in the second interval (of four vibrations) the trolley travelled 1.39 in. farther than it did in the first, and in the third interval it travelled 1.39 in. farther than it did in the second. Apparently²

¹ As the trolley is accelerating the tension in the string is actually a little less than 10 gm. (cf. Ex. 3 on p. 34). The difference, however, is too slight to make any practical difference to the experimental results.

² I.e. as far as our measurements go; but the fact that we get 1.39 in. over might be due to coincidence. By making a greater number of measurements we can show that the acceleration really *is* uniform, and incidentally this establishes the important point that if a given mass is in motion under the action of a constant force, its acceleration will be uniform.

the acceleration is uniform, and we may represent it by the number 1.39.¹

We now treat Fig. 3/2c in the same way (the portion beyond C is not completed in the figure). The acceleration turns out to be 2.78, which is just twice 1.39. Thus doubling the applied force we have also doubled the acceleration. We may try other applied forces, e.g. 25 gm., 30 gm. etc., and make a graph in which applied force (F) is plotted against acceleration (a). It will be found that a straight line is obtained, showing that *the acceleration is proportional to the applied force*.

So much for question (i), in which we altered the force but kept the mass constant. Let us now consider question (ii), in which we alter the mass but keep the force constant.

As already mentioned, there are four heavy steel cylinders of similar shape and size, built into the trolley, and by taking out one or more of these we can readily alter the mass. For instance, suppose the trolley weighs 1600 gm., and that each of the cylinders weighs 400 gm. By experiments similar to that already described, we find the acceleration *using constant moving force* (say 15 gm.) for (i) trolley alone, 1.39 gm., (ii) trolley + one cylinder, total 2000 gm., (iii) trolley + two cylinders, total 2400 gm. and so on, up to trolley + four cylinders, total 3200 gm.

(N.B. With each increase in the load, the friction will increase. We correct this source of error by slightly increasing the weight in the box, and then adjusting the slope, so that the trolley shall move with uniform velocity. This adjustment must be made for each of the loads with which we experiment.)

On working out our results we find that *the acceleration is inversely proportional to the weight of the trolley* (including from none to four cylinders).

Now, it cannot really be the *weight* of the trolley that is affecting the acceleration, for the weight acts vertically

¹ 1.39 is not the actual acceleration, which would be velocity added per second, while 1.39 is the velocity added in the time occupied by 4 vibrations—not necessarily a second. Still, for comparative purposes the time occupied by 4 vibrations will serve us just as well as 1 sec., so long as we stick to this same unit of time throughout.

wnwards, and so could not affect movement in a horizontal section. But if the quantity concerned is not weight, it must be something proportional to weight. This something is mass. Hence we conclude that *for a constant moving force, the acceleration is inversely proportional to the mass moved, or*

$$\text{mass} \times \text{acceleration} = \text{a constant.}$$

Mathematical Expression. We have seen that (i) when the mass is constant, *acceleration* \propto *applied force*, and (ii) when the force is constant, *mass* \times *acceleration* is constant. Both these rules are included in the statement $F \propto ma$, where a is the acceleration produced in a body of mass m by a force F . It is important to notice that ma is equal to the *rate of change of momentum*. For suppose a body of mass m has velocity u , which in the course of a very short time t changes to v . The momentum of the body has changed from mu to mv , and as this has taken place in time t , the rate of change of momentum is $\frac{mv - mu}{t}$

$$\begin{aligned} &= m \times \frac{v - u}{t} \\ &= m \times (\text{rate of change of velocity}) \\ &= m \times \text{acceleration} \\ &= ma. \end{aligned}$$

Thus $F \propto ma$ or $ma \propto F$ is equivalent to saying that the rate of change of momentum is proportional to the applied force. This is Newton's *Second Law of Motion*, which may be formally stated thus: *Rate of change of momentum is proportional to the applied force, and takes place in the direction in which that force acts.*

Absolute Unit of Force. It is shown in mathematics that $F \propto ma$, then $F = kma$, where k is some constant. Now, we define our unit of force as that force which, when acting upon unit mass, causes it to move with unit acceleration, we have F , m and a all $= 1$ in our equation, which then becomes $1 = k \times 1 \times 1$. Thus $k = 1$, and our equation now reads $F = ma$. This equation is of great importance, and is often referred to as the *fundamental kinetic equation*. Notice that it holds good only when the unit of force is defined the way just stated.

Impulse of a Force. We can express Newton's second law in such a way as to bring out more clearly the question of *time*. For as we saw on the preceding page, $\frac{mv - mu}{t} = a$ and $ma = F$,

$$\therefore F = \frac{mv - mu}{t}, \text{ or } Ft = mv - mu.$$

Thus the change of momentum, $mv - mu$, produced in a body of mass m by a force F acting for time t , is equal to Ft . This product (*force*) \times (*time*) is known as the **impulse** or **the force**.

It often happens that a body of considerable mass has its velocity very rapidly reduced. When we hit a nail with a hammer, for instance, the velocity, u , is very quickly brought to zero. Hence $Ft = mv - mu = 0 - mu = -mu$ (the minus sign indicating that while the original velocity is downward, the change in velocity is upward). As m is considerable, so also is Ft . But t is very small, and therefore F is very large. Thus great force is brought to bear upon a nail.

When making a catch at cricket, a player experiences the impulse Ft . By drawing back his hands as the ball reaches him, he greatly increases t and correspondingly reduces F .

Absolute unit and gravitational unit. In previous work we have very often used another unit of force—the pound weight. This is called a *gravitational* unit, because its magnitude depends on the force of gravity. For instance, in Chapter IV of our earlier book we saw that 1 lb. wt.—the force with which a mass of 1 lb. is attracted to the earth—is slightly greater at the north pole than at the equator.

On the other hand, our absolute unit is quite independent of gravity. Imagine a smooth horizontal table in which is a long straight groove. In the groove there rests a smooth platinum ball whose mass is 1 lb. (i.e. there is the same amount of platinum in it as in the standard pound). To this mass we now apply a steady push, enough to make it move with an acceleration of 1 foot per sec. per sec. This steady push is, of course, our *absolute* unit, commonly known as the *dyne*.

poundal. Because the movement is horizontal, the force required is quite independent of gravity. If it were possible to carry out our 'horizontal table' experiment on the surface of the moon (where the force of gravity is only about $\frac{1}{6}$ th of what it is on the earth), the force required to produce unit acceleration would still be exactly the same. 'Absolute' means 'loosed from'—loosed or released from any dependence on the force of gravity.

Relation between the two units. When a mass of 1 lb. is falling freely, it is being pulled down by its own weight—1 lb. wt.—and this force causes it to move with an acceleration of g ft. per sec. per sec. (where g = about 32). Hence we may write

1 poundal acting on a mass of 1 lb. makes it move with acceleration 1

but

1 lb. wt. acting on a mass of 1 lb., makes it move with acceleration g .

$$\therefore 1 \text{ lb. wt.} = g \text{ poundals.}$$

$$\text{or } 1 \text{ poundal} = 1/g \text{ lb. wt.}$$

$$= \frac{1}{32} \text{ ,, ,, (approx.)}$$

$$= \frac{1}{2} \text{ oz. weight approx.}$$

The dyne. In the c.g.s. system our unit of mass is not 1 lb., but 1 gm., and our unit of acceleration is an acceleration of 1 centimetre (not 1 ft.) per sec.²

Thus the absolute unit of force in this system is *that force which when acting upon a mass of 1 gm. causes it to move with an acceleration of 1 centimetre per second per second.* This unit of force is known as the **dyne**.

Just as 1 poundal = $1/g$ lb. wt., so 1 dyne = $1/g$ gm. wt. Notice, however, that while in the first case g = approx. 32 ft. per sec. per sec.) in the second case g = approx. 981 cm. per sec. per sec). Thus 1 dyne = $1/981$ gm. wt., and so it is approximately equal to the weight of 1 milligram.

We will now consider some numerical examples relating to the subject matter of the preceding pages.

Example 4. With good brakes, a motor-car travelling at 30 m.p.h. can be brought to rest in a distance of 30 ft. Show that the total distance to its motion must be approximately equal to its own weight (take $g = 32.2$).

Let us first find the acceleration (negative).

30 m.p.h. = 44 ft./sec.

$$\therefore u = 44. \quad v = 0. \quad s = 30. \quad a = ?$$

$$\text{From } v^2 = u^2 + 2as$$

$$\text{we have } 0 = 44^2 + 60a.$$

$$\therefore a = -\frac{484}{15} \text{ ft./sec.}^2$$

Suppose the car weighs m lb.

Then from $F = ma$, we have force resisting motion = $m \times \frac{484}{15}$ pounds.

$$= \frac{m \times 484}{15 \times 32.2} \text{ lb.wt.}$$

$$= 1.002 m \text{ lb.wt.}$$

Thus the force resisting motion is almost exactly equal to m lb., the weight of the car.

Questions

1. Describe Fletcher's Trolley, and explain how you would use it to test the statement 'force is proportional to mass \times acceleration.'

The lengths of six successive individual waves on a Fletcher's trolley trace are 3.2 cm., 3.5 cm., 3.8 cm., 4.1 cm., 4.4 cm., and 4.7 cm. The period of the vibrator is one-fifth of a second. Find (a) the acceleration of the trolley, and (b) the average velocity over the whole period covered by these readings. *Oxf.*

2. State Newton's second law of motion, and show how it can be used to define a unit of force.

A balloon and contents of total mass 2000 lb. are falling freely with an acceleration of 2 ft. per sec.² Find the change in the vertical motion after 125 lb. of ballast have been thrown out. Neglect air friction and take $g = 32$ ft. per sec.² *O.C.*

3. A barrel weighing 100 lb. is being pulled up a smooth plane inclined at 30° to the horizontal by means of a rope which is parallel to the surface of the plane. If the tension in the rope is 130 lb. wt., what is the acceleration of the barrel? *O.C.*

4. A train of mass 200 tons travels at 60 m.p.h. The brakes are suddenly applied, producing a retarding force equal to one-tenth the weight of the train. Find the distance the train runs on a level track before coming to rest. *O.C.*

CHAPTER 4

NEWTON'S THIRD LAW OF MOTION

If a man presses his thumb against a table, the table presses with an equal force against his thumb—otherwise his thumb would go through the table. If a load is attached to a spring balance and exerts a downward force of 3 lb. wt. on the spring, the spring exerts an upward force of 3 lb. wt. on the load. A magnet attracts a piece of iron with a certain force (not necessarily enough to make it move). The iron attracts the magnet with an equal force, but in the opposite direction.

The cases just noted, and thousands of others (a few of which we will consider presently), are summed up in Newton's **Third Law of Motion**. This states that *to every action there is an equal and opposite reaction*.

From the examples given, it might be thought that the law applies only to bodies at rest, but this is by no means the case. Let us see how it applies to the very common action of walking. Here the force which gives us our 'action and reaction' is *friction*, this being the force which prevents one surface from sliding over another (or tends to prevent it). When a man is walking, his foot pushes backwards against the ground, and the ground pushes forward against his foot. It is the action of this forward force on the man's body that causes him to advance. On a smooth surface such as ice the frictional forces just mentioned are small, and the man finds it difficult to move forward.

Sledge. Horse and Cart. If a man is pulling a sledge forward with a force T_1 (Fig. 4/1), the sledge is pulling him backwards with an equal force T_2 . How, then, is it possible for any advance to be made?

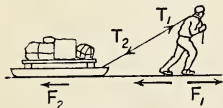


FIG. 4/1

Let us think of the man-plus-sledge as a *whole*. Suppose the force of friction called into play, as the man braces his feet against the ground.

is F_1 . He pushes backward with a force F_1 , but probably the only effect is to disturb a little snow. The important

point is that the ground pushes the man forward with a force F_1 .

Suppose that when the sledge begins to move, it has to overcome a frictional force F_2 . Then if F_1 is greater than F_2 the man-plus-sledge will advance under the action of a resultant forward force $F_1 - F_2$. When we come to the more detailed study of friction (Chapter 9), we shall see that F_2 cannot be greater than F_1 , but it may become equal to it. In that case $F_1 - F_2 = 0$, and the man cannot move the sledge. It will be seen that T_1 and T_2 are internal forces, and we are not really concerned with them any more than we are with the internal forces which hold the various parts of the sledge together.

The case of a horse pulling a cart is very similar. F_1 represents the total frictional forces between the horse's feet and the ground, and F_2 the 'rolling' frictional forces between the two wheels and the ground. The tension in the traces consists of internal forces represented by T_1 , T_2 ($T_1 = T_2$, but acts in the opposite direction).

When considering matters such as walking, a horse pulling a cart, etc., we pay little attention to the backward push of the man's feet (or horse's hoofs) on the earth, because the latter is much too big to be moved appreciably. Often, however, this backward force is exerted against an object of only moderate size, and then we may have to consider how this object is affected. We have a good example in the case of a shot fired from a gun.

Recoil of a gun. Suppose M and m are the masses of gun and bullet respectively. Each is acted on by the explosive force for the same time t . At the end of this time, suppose the gun has acquired a velocity V and the bullet a velocity v . Evidently the gun has acquired a momentum MV and the bullet a momentum mv . The *rates* of change of momentum

$\frac{MV}{t}$ and $\frac{mv}{t}$ respectively, t being the same for each.

But as the forces acting on gun and bullet are exactly equal (Newton's third law), the rates of change of momentum of the latter are also equal (second law), i.e. $MV/t = mv/t$.
 $MV = mv$.

Thus the two bodies subject to the action and reaction acquire equal momenta (in opposite directions). This result applies not only to the particular case discussed—that of the gun and bullet—but also to all cases of action and reaction.

If $MV = mv$, it follows that $M/m = v/V$. Thus we have a means of comparing two masses. Let them 'act and react' upon one another, and note the velocity acquired by each. The respective masses are in the inverse ratio of these velocities.

Conversely—and this is of much more practical interest—if we know the ratio of the masses, we can calculate that of the velocities. Let us consider an example.

Example 1. *A military rifle weighing 9 lb. fires a bullet weighing 1 oz. with a muzzle velocity of 2000 ft./sec. With what velocity will the gun begin to recoil?*

Let required velocity = V ft./sec.

Then momentum acquired by gun = $9 \times V$ f.p.s. units.

and momentum acquired by bullet = $\frac{1}{16} \times 2000$ f.p.s. units.

Equating, we have $9V = \frac{2000}{16}$

$$\therefore V = \frac{2000}{16 \times 9} = 14 \text{ ft./sec. approx.}$$

Similar cases. In a Catherine wheel the 'gun' is free to move in a circular direction, and the effect is too well known to need description. If a man in a boat A is pushing another boat B with an oar, A (if free to move) acquires a momentum equal to that of B, but in the opposite direction. If A is a barge and B a small rowing-boat, the velocity of A may be too small to be perceptible, though its momentum will still be equal to that of B.

Suppose you are standing on a plank that bridges a small stream, and you jump into the air, the effect on the plank is obvious enough—it acquires a downward momentum equal to your own momentum upwards. If when you make your upward jump you are standing on solid ground, it is still true that the earth acquires an equal downward momentum. However, taking your own weight as 1 cwt., and that of the earth as 6×10^{21} tons, it is evident that the velocity acquired by the earth will be of no very disturbing character.

Questions

1. State Newton's Laws of Motion.

A wooden block, of mass 5 lb., is pulled along a rough horizontal table by means of a string parallel to the surface of the table, which passes over a light, frictionless pulley at the edge of the table and carries a weight of 3 lb. The resistance offered by the surface of the table to the motion of the block is 2 lb. wt. Calculate the acceleration with which the block moves. ($g = 32 \text{ ft./sec.}^2$, *Oxf.*)

2. State the principle of *conservation of momentum* and show how it follows from Newton's laws of motion.

If a force of 10 lb. wt. acts steadily on a mass of 1 ton, how long will elapse before the mass has moved 14 ft. if it starts from rest? *O.C.*

3. A spring balance in a lift supports a body of mass m grams. Describe and account for any changes which occur in the spring balance reading (*a*) when the lift is rising or falling with uniform velocity, (*b*) when falling with uniform acceleration, (*c*) when rising with uniform acceleration. *Dur.*

4. 'To every action there is an equal and opposite reaction.' What is meant by this statement? Use it to explain the changes in the horizontal motion of a dinghy when a man in the bows walks to the stern, stands motionless, and then dives into the water.

A dinghy, originally at rest, is pulled for 3 sec. with a steady force of 25 lb. weight. Calculate its velocity at the end of this time and the work done in pulling it, assuming the water resistance to be negligible. (Mass of dinghy = 600 lb., $g = 32 \text{ ft./sec. per sec.}$) *J.M.B.*

5. Define (*a*) velocity, (*b*) momentum of a moving body. What is the relationship between the force applied to a moving body and the momentum which it imparts to the body?

A force P acts for t seconds on a body of mass m , which is at rest, and gives it a velocity v . The same force P then acts for T seconds on another body of mass M at rest and gives it a velocity V .

Find the ratio m/M in terms of t , v , T and V .

Distinguish between the mass of a body and its weight. *Camb.*

CHAPTER 5

WORK, ENERGY AND POWER

WE have more than once had occasion to recall the definition of Force as 'that which changes or tends to change the state of rest or motion of a body.' Now, when a force merely *tends* to change a body's state of rest, it may be very useless from a practical point of view. A furniture remover who found that one of his men had been pushing and pulling at a grand piano for half an hour, without moving it an inch would be more likely to dismiss him for stupidity than to commend him for perseverance. 'Force is all very well,' he would say, 'but only when it produces movement.'

Force when it produces movement is said to *do work*. Examples abound on every hand—a horse pulling a cart, steam driving a piston, an electromagnet picking up scrap iron, and so on.

Sometimes movement takes place in a direction opposite to that in which the force is acting. A horse beginning to pull a load up a hill may find himself dragged backwards, especially if the ground be slippery. In such a case the work done by the horse is negative, but we usually say that work is done 'against the force.' If you lift a book or a chair, you do work against the force of gravity.

We shall now have no difficulty in understanding a formal definition.—*A force is said to do work when it produces movement in the direction of its line of action; if movement takes place in the opposite direction, work is said to be done against the force.*

We must now consider how work can be measured, i.e. we must fix upon a unit. Evidently this will have to take into account (i) the force itself, (ii) the distance through which it has caused movement to take place. One very common unit is the **foot-pound**, which is *the work done when a force of 1 lb. wt. moves its point of application through a distance of 1 ft.* We do 1 ft.lb. of work when we raise a pound weight through a vertical height of 1 ft. If we raise it 3 ft., we do

ft.lb. To raise a 5 lb. weight through a vertical height of ft. we have to do $5 \times 4 = 20$ ft.lb. and so on.

Of course the direction in which work is done is by no means always vertical. Suppose the 5 lb. weight just mentioned were being dragged along a table, and the dragging force required were 2 lb. wt., then the work done in dragging the weight 4 ft. would be $2 \times 4 = 8$ ft.lb. In this case the work done against the force of friction.

Example 1. A labourer weighing 10 stones carries a pile of slates weighing 50 lb. to the top of a house 30 ft. high. Find (i) the total work done, and (ii) the useful work.

Weight of man = 10 stones = 140 lb.

„ „ slates „ = 50 lb.

Total weight = 190 lb.

\therefore total work = $190 \times 30 = 5700$ ft.lb.

The only *useful* work is that done in lifting the slates,
 \therefore useful work = $50 \times 30 = 1500$ ft.lb.

Other units of work. The foot-pound is of course a *gravitation* unit, and as the pound-weight varies slightly from place to place the foot-pound varies correspondingly.

An *absolute* unit of work would be the **foot-poundal**—the work done when a force of 1 poundal moves its point of application through a distance of 1 ft. The foot-poundal = $\frac{1}{g}$

lb., where $g =$ approx. 32 (ft. per sec. per sec.).

In the c.g.s. system we have as a gravitation unit the **centimetre-gram**, a term which explains itself. The corresponding absolute unit would be the work done when a force of 1 dyne (p. 33) moves its point of application through a distance of 1 centimetre. This unit is known as the *erg*.

The erg = $\frac{1}{g}$ centimetre-grams, where $g =$ approx. 981

(cm. per sec. per sec.). Thus the erg would be roughly equal to the work required to lift a milligram weight through a vertical height of 1 cm. This unit is so very small that we often use the **joule** as our absolute unit, the joule being equal to 10^7 ergs.

Power. If a man is buying a machine to do work—perhaps a pump to raise water from a mine, or a crane to raise blocks

of stone, etc., during building operations—he does not ask how much work the machine can do, but the *rate* at which it can do it. Rate of working is known as *power*.

The obvious unit of power seems to be a rate of working of 1 ft.lb. per second. For practical purposes, however, this is much too small, and a unit known as the **horse-power** is employed, this being *a rate of working of 550 ft.lb. per second*.

This unit is due to the famous engineer James Watt. He arranged to have a barge pulled along a canal by a good dray-horse, a strong spring-balance being fixed in the tow-rope. He multiplied the force in pounds weight by the speed of the horse in feet per second, and so obtained the number of foot-pounds done per second. After making certain corrections he came to the conclusion that a horse can work at the rate of 550 ft.lb. per second. Actually, an average horse cannot maintain this rate of working for any very long period.

Example 2. *A man weighing 180 lb. bounds up a flight of stairs (12 steps each 8 in. high) in 3 sec. At what horse-power is he working?*

Total height = $12 \times 8 \text{ ins.} = 8 \text{ ft.}$

Work done in lifting his own weight (180 lb.) through a vertical height of 8 ft. = $180 \times 8 \text{ ft.lb.}$

He does this in 3 sec.

$$\therefore \text{work done per sec.} = \frac{180 \times 8}{3} \text{ ft.lb.}$$

$$\therefore \text{H.P.} = \frac{180 \times 8}{3 \times 550} = 0.87.$$

(N.B. Of course a man could work at this rate only for a very short time.)

Example 3. *'The mean flow of the Niagara river is about 222,000 cub. ft. per second with a fall of 160 ft.'* *Encyc. Brit.*

Calculate the available horse-power (1 cub. ft. of water = 62.4 lb)

Water falling per sec. = $222,000 \times 62.4 \text{ lb.}$

Fall = 160 ft.

$$\therefore \text{Work done per sec.} = 222,000 \times 62.4 \times 160 \text{ ft.lb.}$$

$$\therefore \text{H.P.} = \frac{222,000 \times 62.4 \times 160}{550}$$

$$= 4,030,000 \text{ approx.}$$

In the examples just given the work has been done by, or against, gravity, and such cases are commonly met with in daily life. Even commoner, however, are cases in which

work is done against friction. Think, for instance, of the work done by a horse pulling a plough, or by a man cycling on a level road.

Example 4. A train weighing 100 tons travels on the level at a uniform speed of 45 m.p.h. The frictional resistances are equivalent to 10 lb. weight per ton. Find the H.P. at which the engine is working.

The frictional resistances are equal to a backward force of $10 \times 100 = 1000$ lb.wt.

Now since the train is moving with uniform velocity, the resultant force on it is zero (by Newton's second law). Therefore the engine must be exerting a forward pull of 1000 lb.wt. to balance the frictional forces.

$$45 \text{ m.p.h.} = 66 \text{ ft. per sec.}$$

$$\therefore \text{work done per sec.} = 1000 \times 66 \text{ ft.lb.}$$

$$\therefore \text{H.P.} = \frac{1000 \times 66}{550} = 120.$$

Energy. An energetic man is one who habitually does a great deal of work. If he is merely *capable* of doing work, but does not do it, we describe him not as energetic, but as *lazy*.

In science, however, a body is said to possess energy if it is capable of doing work. Moving air (i.e. wind) possesses energy, for it is capable of driving a ship along, or turning a windmill. A shell fired from a gun can do the (destructive) work of knocking down a bridge or a house. When a grandfather clock has just been wound up, the weights possess energy, because as they slowly descend they will do the work of moving the various parts of the clock. If we stop the clock, we still say that the weights possess energy, because they are *capable* of doing work though they are not actually doing it.

When we examine more closely the cases we have just noted, we see that they are of two different kinds. The moving air and the shell are capable of doing work because they are *moving*; the weights of the clock are capable of doing work because of their high *position*. These two kinds of energy are known respectively as *kinetic* and *potential*, i.e. a body possesses **kinetic energy** when it is capable of doing work on account of its motion; it possesses **potential energy** when it is capable of doing work on account of its position.

Potential energy may be turned into kinetic, and vice versa. Of this we have a good example in the swinging of a pendulum. Suppose the bob is at A on the extreme left of its path (Fig. 5/1). It is in its highest position and can do work as it descends the path AV, i.e. it has potential energy. On the other hand, it is for a moment stationary while in the position A, and so possesses no kinetic energy.

Now let the bob swing into the position V. It has the reached its lowest point, and so has lost all its potential energy. But it has attained its maximum speed, and so its kinetic energy is at a maximum. In short, at A potential energy but no kinetic; at V kinetic energy but no potential. The one kind has given place to the other. As the bob ascends the arc VB, its kinetic energy will decrease, sinking to zero at B, where its potential energy will once more have reached a maximum.



FIG. 5/1

This conversion of one kind of energy into the other is also seen in the case of the pile-driver. Here a heavy iron block is hauled up to a position near the top of a high tower of scaffolding, i.e. potential energy is imparted to it. It is then allowed to fall, acquiring kinetic energy as it does so, and this kinetic energy is made to do the useful work of driving in the piles. The processes of hauling up and then releasing are of course repeated as often as required.

Measurement of energy. As the energy of a body is its capacity for doing work, we naturally measure this energy by the work the body is capable of doing—in foot-pounds or foot-poundsals, for instance. Thus if the 'weight' of a grandfather clock has a mass of 8 lb. and has been raised to a height of 3 ft., its potential energy is $8 \times 3 = 24$ ft.lb. For imagine this mass A to be attached to another equal mass B by a thread passing over a pulley. Then if there were no friction or air resistance, A, in falling 3 ft., would be just able to raise B 3 ft., i.e. the work it is capable of doing—its energy—is 24 ft.lb. 24 g ft.pdls. Speaking generally, we see that if a mass



B

FIG. 5/2

a lb. is at a height of h ft. above the lowest position it can occupy, then its potential energy is mh ft.lb., or $mh \times g$ p.dls.

So much for potential energy. With regard to kinetic energy, it can be proved that the kinetic energy possessed by a body of mass m moving with velocity u , is equal to $\frac{1}{2} mu^2$.

For suppose the body is brought to rest by a force F , after travelling a distance s . Let a be the acceleration (negative).

Applying the formula $v^2 = u^2 + 2as$ we have $v = 0$, so that $0 = u^2 + 2as$.

$$\therefore a = -\frac{u^2}{2s}.$$

The work done by the force F , acting through a distance $s = Fs$.

But $F = ma$ (p. 31).

\therefore work done by force $F = mas$

$$= m \times \frac{-u^2}{2s} \times s$$

$$= -\frac{1}{2} mu^2.$$

kinetic energy of moving body

= work done *against* force F in process of coming to rest

$$= -(-\frac{1}{2} mu^2) = \frac{1}{2} mu^2.$$

Example 5. What is the kinetic energy of a mass of 6 lb. moving with a velocity of 10 ft. per sec.?

$$\text{Required K.E.} = \frac{1}{2} mu^2$$

$$= \frac{1}{2} \times 6 \times 10^2$$

$$= 300 \text{ ft.pdls.}$$

Energy and Momentum. We must take care not to confuse these two terms. We have just seen that the kinetic energy of a mass m moving with velocity u is $\frac{1}{2} mu^2$, while its momentum is equal to mu (p. 27).

Notice that if the velocity of a body is trebled, its momentum is also trebled, but its energy is increased nine-fold. Similarly, of course, for other numbers. Thus we should expect a body moving with very high velocity to be remarkable for its energy rather than for its momentum.

This is clearly seen in the example overleaf.

Example 6. *A military rifle weighing 9 lb. fires a bullet weighing 1 oz. with a muzzle velocity of 2000 ft./sec. Find (i) the momentum (ii) the energy of both gun and bullet just after the shot is fired.*

This example was studied on p. 38 from the point of view of momentum. There it was shown that the momentum of the bullet is equal to $\frac{1}{16} \times 2000$ or 125 f.p.s. units; and it follows from Newton's 2nd and 3rd laws that the momentum of the gun is equal and opposite to this.

It was also shown that the gun begins to recoil with a velocity of 14 ft./sec.

$$\therefore \text{K.E. of gun} = \frac{1}{2}mv^2 = \frac{1}{2} \times 9 \times 14^2 = 882 \text{ ft.pdls.}$$

$$\begin{aligned} \text{But K.E. of bullet} &= \frac{1}{2} \times \frac{1}{16} \times 2000^2 \\ &= 125,000 \text{ ft.pdls.} \end{aligned}$$

Thus the kinetic energy of the bullet is enormously greater than that of the gun, though the two have equal momenta.

Equivalence of P.E. and K.E. On p. 44 we considered examples in which a loss of potential energy is accompanied by a gain in kinetic energy, and vice versa. This at once suggests the question, 'when a body loses potential energy, does it gain an equivalent amount of kinetic energy, and vice versa?'

The following example will throw much light on the question.

* **Example 7.** *A body of mass 10 lb. is raised to a vertical height of 80 ft. Find the total of its P.E. and K.E. (i) at this height, (ii) after it has fallen 40 ft., (iii) after it has fallen to ground level.*

(i) Before being allowed to fall its P.E. = $10 \times 80 = 800$ ft.lb. or $800 \times 32 = 25,600$ ft.pdls.

$$\text{K.E.} = 0$$

$$\therefore \text{P.E.} + \text{K.E.} = 25,600 \text{ ft.pdls.}$$

(ii) After falling 40 ft. its P.E. is halved and is $\therefore 12,800$ ft.pdls. To find its K.E. we must first calculate its velocity. We have $u = 0$, $s = 40$, $a = 32$, $v = ?$

Applying the equation $v^2 = u^2 + 2as$ this gives $v^2 = 0 + 2 \times 32 \times 40 = 2560$.

$$\begin{aligned} \therefore \text{K.E.} &= \frac{1}{2}mv^2 = \frac{1}{2} \times 10 \times 2560 \\ &= 12,800 \text{ ft.pdls.} \end{aligned}$$

$$\begin{aligned} \therefore \text{P.E.} + \text{K.E.} &= 12,800 + 12,800 \\ &= 25,600 \text{ ft.pdls.} \end{aligned}$$

(iii) After falling to ground level, P.E. = 0.

To find its velocity, we have $u = 0$, $s = 80$, $a = 32$, $v = ?$

$$\therefore \text{from } v^2 = u^2 + 2as$$

$$\begin{aligned}\text{we get } v^2 &= 0 + 2 \times 32 \times 80 \\ &= 5120\end{aligned}$$

$$\begin{aligned}\therefore \text{K.E.} &= \frac{1}{2}mv^2 = \frac{1}{2} \times 10 \times 5120 \\ &= 25,600 \text{ ft.pdls.}\end{aligned}$$

$$\therefore \text{P.E.} + \text{K.E.} = 25,600 \text{ ft.pdls.}$$

Thus we see that the *combined* P.E. and K.E. of the body the same in all three cases, and it can easily be shown that would be the same in any other position. As it falls, it loses potential energy, but gains an *equal amount* of kinetic energy.

The foregoing is, of course, only a particular example of general case. The latter is not at all difficult to prove, but here we will content ourselves with merely stating it.—a body of mass m falls from a height h , the sum of its P.E. and K.E. is constant and equal to mgh .

It is assumed in this statement that the resistance of the air may be neglected. In practice, some air is set in motion as the body falls, and a more exact statement would be that loss of P.E. of falling body = gain of K.E. of body *plus* gain of K.E. of air.

It can be shown that the same principle holds good for a body running down a slope, i.e. loss of P.E. = gain of K.E. In this statement we are neglecting not only the resistance of the air, but the frictional resistance of the surface of the slope.

Conservation of energy. We have seen that as a body falls it loses its potential energy and acquires a corresponding amount of kinetic energy. But what happens when it reaches the ground? It has lost all its potential energy, and as it comes to a stop it has presumably lost all its kinetic energy.

We can get an idea of what really happens if we let our

'falling body' consist of a closed picnic-basket packed with crockery. This has unfortunately slipped over the vertical edge of a quarry.

On striking the ground the kinetic energy of the body as a whole disappears, but the constituent parts, the crockery at once begin to shake about—let us use the word *vibrate*—with great violence (with woeful results). We can trace what has happened to some at least of the original kinetic energy. It has been transferred from the whole to the constituent parts.

But suppose the falling body is of a more tightly packed character—a lead bullet, for instance? In this case the constituent parts are the molecules. There is good reason to believe that these are always in a state of vibration, and when a moving bullet comes to a sudden stop, the rate of vibration of the molecules is greatly increased. This increased rate of vibration makes itself felt as *heat*. Indeed, a bullet that has just been brought to a sudden stop by striking an iron target is usually too hot to handle.

There are abundant examples of this conversion of kinetic energy into heat. When a bicycle pump is worked it soon becomes warm. A match rubbed against a rough surface becomes hot enough to take fire. Saws, files, drills, and so on, become hot in use; and so on.

Now, when considering potential and kinetic energy, we saw that one could be converted into an exact equivalent of the other. It is tempting to enquire whether 'mechanical' energy can be converted into an exact equivalent of heat.

Actually it can. It was discovered in 1847 by Dr. Joule of Manchester that 774 ft.lb. are equivalent to a unit of heat now called the British Thermal Unit¹ as definitely as 12 pence are equivalent to a shilling. Are there still other forms of energy? Is it possible that no energy is ever really lost or created, though it may be turned from one form into another? These are questions which we shall be better able to answer when we have made some progress in other branches of Physics.

¹ I.e. the amount of heat required to raise the temperature of 1 lb. of water 1 degree Fahrenheit.

Questions

A

1. How much work is done when
 - (i) a 7 lb. weight is lifted on to a table 3 ft. high,
 - (ii) a man weighing 10 stones climbs a ladder 12 ft. high,
 - (iii) a chair weighing 20 lb. is dragged 5 ft. across a floor, average frictional resistance being 3 lb.wt.?
2. Find the horse-power of a pump which every minute raises 5 galls. of water (1 gall. = 10 lb.) to a height of 40 ft.
3. In one hour a pump lifts 20,000 gallons of water 30 ft. above the level of the surface of a reservoir. What horse-power is required for the pump? (1 gall. = 10 lb.).
4. Find
 - (i) the potential energy possessed by a clock weight of 3 lb. which has been raised to a height of 5 ft. above its lowest point.
 - (ii) the kinetic energy of a bullet weighing 1 oz. and travelling at the rate of 1000 ft. per sec.
 In both cases express your answer in ft.pdls.

B

1. Calculate the two quantities *momentum* and *kinetic energy* of a body of mass 2 lb., dropped from height, (i) when it reaches the ground 81 ft. below, (ii) when it has reached halfway in its fall. *C.W.B.*
2. What is meant by *force*, *work*, *power*?
A cyclist rides up a gradient of 1 in 44 (measured along the slope) at the rate of 6 miles per hour. If the total mass of man and machine is 165 lb., calculate the horse-power he is developing. Why would your calculated value of the horse-power be lower than the real horse-power necessary? (1 H.P. = 33,000 ft.lb. wt. per minute). *Lond.*
3. A man weighing 12 stones runs up a flight of steps and in 10 sec. is 40 ft. above ground level. How much work does he do and at what average rate in horse-power is he working? *Dur.*
4. Explain the terms 1 lb.wt., 1 dyne, 1 ft.lb., a horse-power. What force acting along a smooth plane inclined at 45° to the horizontal will keep a mass of 10 lb. on the plane in equilibrium?
A cyclist rides at 12 m.p.h. along a level road. If he expends energy at the rate of 0.1 horse-power, what is the average value of the frictional forces? *Lond.*
5. If a pressure of 25 lb.wt. per sq. in. acts on the circular piston (radius 10 in.) of a steam engine, and drives the piston through a distance of 5 ft., find the work done in ft.lb. weight. *Cumb.**

6. A train and its engine, weighing together 200 tons, are kept moving on a level track at 40 m.p.h. If the resistance to motion is 30 lb. per ton, what horse-power is being developed by the engine? *Camb.**

7. Define the terms *momentum*, *kinetic energy*.

A motor-car weighing 1 ton is travelling at 30 m.p.h. Find the average force which must be exerted by the brakes to bring it to rest (a) in 11 sec., (b) in 44 yd. *Oxf.*

8. Describe any means by which a man who can exert a force of not more than 220 lb.wt. can increase the potential energy of a body weighing 0.75 ton by 5000 ft.lb.

An aeroplane flying on a straight course at a height of 1000 ft has its engine switched off and glides down to an aerodrome. Describe the changes in energy from the time of switching off till the plane comes to rest on the runway. *J.M.B.*

9. (a) Define (i) the *momentum* of a body, and (ii) its *kinetic energy*. A body of mass 100 gm. falls freely from rest through 490 cm. What is the gain in (a) kinetic energy, (b) momentum ($g = 980 \text{ cm./sec./sec.}$)

(b) A shell fired horizontally from the muzzle of a gun breaks an electric circuit near the muzzle, so causing a plate to fall from rest. Simultaneously it makes a pointer trace a mark A on the plate. After travelling another 100 ft., the shell causes the pointer to trace another mark B on the falling plate, and it is found, by means of a microscope, that the distance AB is 1.51 in. Find the average velocity with which the shell traverses this 100 ft. (Take $g = 32.0 \text{ ft./sec./sec.}$) *Camb.*

10. State Newton's second law of motion, and explain how the c.g.s. unit of force (the dyne) is derived.

Calculate the kinetic energy of a motor-car of mass 1 ton moving at a speed of 30 m.p.h., stating the unit. What is the shortest distance in which the car can be stopped, (a) from 30 m.p.h., (b) from 60 m.p.h., if the maximum retarding force is equal to the weight of the car. *Camb.*

11. Describe a method of measuring the *acceleration due to gravity* (g).

A bullet weighing $\frac{1}{2}$ oz. is fired vertically upwards from the ground and reaches a height of 900 ft. Neglecting air resistance calculate (a) the initial velocity of the bullet, (b) the time that elapses before the bullet reaches the ground again, (c) the energy (potential and kinetic) of the bullet at ground level, at the highest point reached, and half-way up. *C.W.B.*

12. State what you understand by *kinetic* and *potential energy* and give two examples of each. A body of mass 100 gm. falls from rest from a point 1000 cm. above the ground. Calculate the potential energy and the kinetic energy after it has been falling for 1 sec.

Describe how the potential and kinetic energies change during the swinging of a pendulum. *Dur.*

CHAPTER 6

MACHINES AND THEIR EFFICIENCY

In physics the word *machine*, is used in a rather special sense to indicate some contrivance for changing either the point of application, the direction, or the magnitude of a force. More often than not our machine enables us to change two or three at once. Thus when a man is winding up a bucket of water from a well, his force is applied at a point yards away from the bucket itself. Again, while the direct force needed to lift the bucket would be an upward one, the man's hand may actually be moving downwards. Lastly, the man may be exerting a force considerably smaller than the weight of the bucket and water.

You are certain to have some knowledge already of a number of machines—the lever, wheel and axle, inclined plane and various systems of pulleys—and you will have learnt how to calculate the *force ratio* (or *mechanical advantage*) and *velocity ratio* in each case. In the present chapter we shall consider machines from a somewhat different point of view—that of the *work* we can obtain from them. At the same time we shall seize the opportunity of making a fairly thorough revision of our previous work, with additions where necessary.

We will begin by recalling some of the chief properties of pulleys.

1. In Fig. 6/1 we have the single fixed pulley, in 6/2 a block and tackle (or 'single-string' system), and in 6/3 a 'separate-string' system.

As we shall almost immediately have occasion to use the terms 'force ratio' (or 'mechanical advantage') and 'velocity ratio,' let us be sure that we remember their meaning.

The *force ratio* or *mechanical advantage* of a machine is

$$\frac{\text{resistance overcome}}{\text{effort applied}}$$

In this definition the 'resistance overcome' does not include

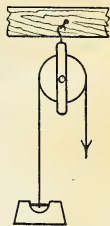


FIG. 6/1



any resistance offered by the machine itself. Thus, referring to Fig. 6/2, suppose the block and tackle is being used to raise a mass of iron L weighing 100 lb., and that the lower block of pulleys weighs 18 lb. If the effort that has to be

applied at E is 30 lb., the force ratio is $\frac{100}{30} = 3\frac{1}{3}$, not $\frac{11}{30}$

because in the latter case we should be including a resistance offered by the machine itself.

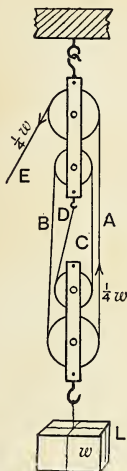


FIG. 6/2

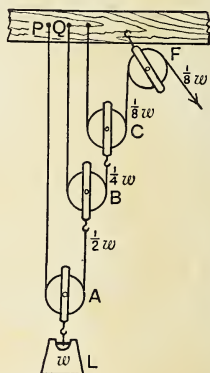


FIG. 6/3

The *velocity ratio* (or *speed ratio*) of a machine is the ratio $\frac{\text{distance through which effort acts}}{\text{distance through which resistance is overcome}}$.

Again referring to Fig. 6/2, if we wanted to raise L 2 feet we should have to pull in 8 ft. of rope, i.e. E would have to move 8 ft. Thus the velocity ratio would be $8 \div 2 = 4$.

We need not spend long over the single fixed pulley Fig. 6/1. If it were frictionless, the effort required to raise

given load would be practically equal to the load. Hence force ratio = resistance \div effort = 1. In practice there is at least a little friction. Hence the effort must be greater than the resistance, and this makes the force ratio less than 1. The velocity ratio is evidently 1, for to raise the load a certain distance, the rope must be pulled through an equal distance.

In Fig. 6/2 suppose the load weighs w lb., and for the moment let us neglect the weight of the lower block, and also the resistance due to friction. The load is supported by four parallel cords, and the strain in each of these is therefore $\frac{1}{4}w$. This would also be the effort required at E. Hence force ratio = $w \div \frac{1}{4}w = 4$. In practice, the effort at E would obviously be greater than $\frac{1}{4}w$, and so the force ratio would be less than 4.

Now for the velocity ratio. Suppose we wish to raise L 1 ft. To do this, each one of the four supporting cords must be shortened by 2 ft., and therefore the length of rope to be pulled in at E is $4 \times 2 = 8$ ft. Thus the velocity ratio is $8 \div 2 = 4$. Notice that neither friction nor the weight of the lower block makes any difference to the velocity ratio—to raise L a certain distance we should still have to pull in four times that length of cord. On the other hand, the force ratio is greatly affected by friction, and by the weight of the lower block.

Fig. 6/3 almost explains itself, but if it presents any difficulty you should refer back to p. 43 of *Junior Physics*. Assuming frictionless and weightless pulleys, the force ratio would be $w \div \frac{1}{8}w = 8$. Further, to raise L 1 ft., it can easily be shown that we should have to pull in 8 ft. of rope at F, and so the velocity ratio is $8 \div 1 = 8$.

Still neglecting friction, let us work out the force ratio, supposing that L weighs 40 lb. and that each pulley weighs 4 lb.

The two parts of the cord A are supporting a total weight of 44 lb., so the strain in each is 22 lb. The two parts of the cord round B are supporting this 22 lb. *plus* 4 lb. (the weight of B itself), total 26, so the strain in each is 13. Similarly, the strain in the next higher cord will be $\frac{1}{2}(13 + 4)$ or $8\frac{1}{2}$ lb., and this will be the effort actually required. Thus the force

ratio will be $40 \div 8\frac{1}{2} = 4.7$ (instead of the ideal 8 with weightless pulleys). The velocity ratio however will still be 8.

Efficiency. Principle of Work. It is evident that pulleys and other machines, enable us to do *work*—a term defined on p. 40. In the case we have just been considering, if we suppose L, weighing 40 lb., to have been raised 1 ft., our system of pulleys has enabled us to do work to the extent of $40 \times 1 = 40$ ft.lb.

But to obtain this result we have had to pull the cord at F through a distance of 8 ft., the force applied being $8\frac{1}{2}$ lb. (see end of last paragraph). In this operation we have done work equal to $8 \times 8\frac{1}{2}$ or 68 ft.lb. We have, so to speak, put 68 ft.lb. *into* our machine and have got only 40 ft.lb. *out* of it. The fraction $\frac{40}{68}$ is known as the *efficiency* of the machine in this particular case. We may say in general that the *efficiency of a machine* is the ratio $\frac{\text{work got out of machine}}{\text{work put into it}}$

Efficiency is often expressed as a percentage. In the case just discussed it would be 59% ($= \frac{40}{68} \times 100$).

The efficiency of a machine may vary a great deal, according to the load. In the case just considered L was taken as 40 lb., each pulley weighing 4 lb. Suppose L had been 80 lb. Then, working in the same way as on p. 53, we easily find that the force required at F would be $13\frac{1}{2}$ lb. Now suppose the 80 lb. load to be raised 1 ft.

$$\text{Work got out of machine} = 80 \times 1 = 80 \text{ ft.lb.}$$

$$\text{Work put into it} = 13\frac{1}{2} \times 8 = 108 \text{ ft.lb.}$$

$$\therefore \text{efficiency} = \frac{80}{108} \times 100 = (\text{approx.}) 75\%.$$

(With a load of 40 lb. we found that the efficiency was only 59%.)

There would be no great difficulty in verifying such results by experiment. Thus we could hang a load of 40 lb. at L and a scale-pan of known weight at F. In this pan we should place weights until a point is reached at which the pan, when released, will just move gently downwards.

We measure the distance of the load, and of the pan, above the table or bench, and then allow the pan and its contained weights to fall as far as convenient, afterwards taking measurements again. We then have

$$(\text{load}) \times (\text{distance risen}) = (\text{work got out}).$$

$$d \quad (\text{weights}) \times (\text{distance fallen}) = (\text{work put in})$$

From these results we readily obtain the efficiency.

Incidentally, we obtain a practical measure of the force ratio by dividing load by effort; and of velocity ratio by dividing distance risen into distance fallen.

What would be the efficiency of our system of pulleys if the pulleys were weightless and moved without friction?

In this case we can easily calculate that to support a load of 40 lb. the force at F would have to be 5 lb. Now raise L 8 ft. The rope at F must be pulled 8 ft. So we have:

$$\text{Work got out of machine} = 40 \times 8 = 320 \text{ ft.lb.}$$

$$\text{Work put into machine} = 5 \times 8 = 40 \text{ ft.lb.}$$

In this case the efficiency would be 100%. In practice, of course, because of friction, etc., such an efficiency can never be quite reached. This is true not only for a system of pulleys but for *any* sort of machine.

While under ideal conditions (never realised in practice) the work got out of a machine would be equal to that put into it, it is certainly true that *the work got out of a machine can never exceed that put into it*. This very important statement is known as the **Principle of Work**. If it were not true, we should be able to create energy, i.e. to violate the principle of Conservation of Energy.

Efficiency, Force Ratio and Velocity Ratio. Suppose that by a system of pulleys, or by any other machine, we are able by means of an effort of w lb.wt., acting through x ft., to raise a load (or overcome a resistance) of l lb. through a distance y ft. By definitions already given,

$$\text{force ratio} = \frac{l}{w}, \text{ and } \text{velocity ratio} = \frac{x}{y}.$$

$$\begin{aligned} \therefore \frac{\text{force ratio}}{\text{velocity ratio}} &= \frac{l}{w} \div \frac{x}{y} = \frac{ly}{wx} \\ &= \frac{\text{work got out of machine}}{\text{work put into machine}} \\ &= \text{efficiency.} \end{aligned}$$

The relation between these three quantities—force ratio, velocity ratio and efficiency—should be carefully remembered.

Example 1. A certain system of pulleys has an efficiency of 7 and a velocity ratio of 8. What force would have to be exerted to lift a weight of 600 lb.? If this weight is lifted 3 ft., calculate work (i) put into the system, (ii) got out of it.

$$\frac{\text{Force ratio}}{\text{Velocity ratio}} = \text{Efficiency.}$$

$$\therefore \frac{\text{Force ratio}}{8} = \frac{3}{4}.$$

$$\therefore \text{Force ratio} = \frac{3}{4} \times 8 = 6.$$

$$\therefore \frac{\text{load}}{\text{effort}} = 6.$$

$$\text{But load} = 600 \text{ lb.wt.}$$

$$\therefore \text{effort} = 100 \text{ lb.wt.}$$

If load is lifted 3 ft., effort acts through a distance of $3 \times 8 = 24$ ft. (because velocity ratio = 8).

$$\therefore \text{work put into machine} = \text{effort} \times \text{distance} \\ = 100 \text{ lb.wt.} \times 24 \text{ ft.} = 2400 \text{ ft.lb.}$$

$$\text{work got out of machine} = \text{load} \times \text{distance} \\ = 600 \text{ lb.wt.} \times 3 \text{ ft.} = 1800 \text{ ft.lb.}$$

The Inclined Plane. On p. 39 of *Junior Physics* we found that if a body of weight W was being drawn up a smooth inclined plane by a force P parallel to the plane (Fig. 6/4) then $\frac{P}{W} = \frac{h}{l}$, where h is the height of the plane and l its length.

It follows at once that $P = \frac{Wh}{l}$.

Let us see how this result squares with the principle of Work.

We may regard our plane as a 'machine' for enabling us to raise the object of weight W from the lower level BC to the higher level A .

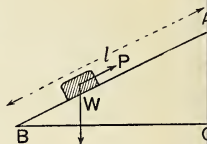


FIG. 6/4

i.e. Work got out of machine = Wh .

Work put into machine = force \times distance

$$= P \times l$$

$$= \frac{Wh}{l} \times l$$

$$= Wh$$

$$= \text{Work got out of machine.}$$

Thus the principle is verified.

The Screw. The screw is a genuine machine in the sense in which we are using the term. In the lifting jack, for instance, is used in raising the great weight of a motor-car.

You will learn a great deal about a screw by cutting out a right-angled triangle ABC, and coiling it round a thin cylinder such as a pencil. This will give you a very fair model of a screw, as shown in Fig. 6/5. Suppose we are driving a screw into wood, or into a steel nut. When we give it one complete turn, it advances by the amount indicated by KB in Fig. 6/5, i.e. the distance from one thread to the next, measured along the axis of the cylinder. This distance is known as the *pitch* of the screw.

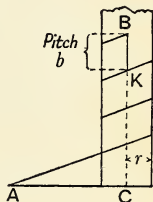


FIG. 6/5

It may conveniently be noted here that the relation between circumference and pitch is made use of in the *micrometer screw gauge*. Suppose, for instance, that the screw has a pitch of 1 mm. We have a horizontal scale (see Fig. 6/6) to show how many complete turns of the screw have been made, and a circumferential scale (CC') to show the number of *hundredths* of a turn. We can then readily take a reading to the hundredth of a millimetre. Thus the ball bearing in the figure has a diameter of 2.54 mm.

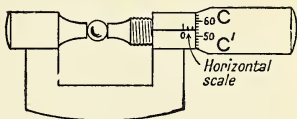


FIG. 6/6

We can find the force ratio

and velocity ratio of the screw by considering it as a coiled inclined plane. It is simpler, however, to make use of the principle of Work, thus:—

Suppose the radius of the screw = r and its pitch = b . Let E = force applied (horizontally, in the direction of a tangent to the cylinder), and R = resistance overcome.

In one complete revolution the force E acts through a distance equal to the circumference, i.e. $2\pi r$. Hence,

$$\text{work put into machine} = E \times 2\pi r \quad \dots (i)$$

At the same time the screw overcomes a resistance R through distance b .

\therefore work got out of machine $= R \times b$. . . (ii)

But in a perfect machine, work put in = work got out.

$$\therefore 2\pi rE = Rb.$$

$$\therefore \frac{R}{E} = \frac{2\pi r}{b}.$$

R/E is of course the force ratio, and it is evident that the smaller we make b , the greater is the force ratio. We can also increase the force ratio by increasing r . For practical purposes we do this by using a screw-driver with a broad handle; or by using a T-piece, as in the lifting jack, or the implement used for turning on a street hydrant.

$$\begin{aligned} \text{Velocity ratio} &= \frac{\text{distance through which effort acts}}{\text{distance through which resistance is overcome}} \\ &= \frac{2\pi r}{b} \\ &= \text{force ratio.} \end{aligned}$$

So far we have assumed a perfect machine, including absence of friction; and velocity ratio = force ratio, as we should expect. In the screw, however, the frictional resistance is very great. Thus the force ratio is much less than the velocity ratio, which is much the same thing as saying that the efficiency is low.

Example 4. A screw-jack is operated by a T-piece having cross-bar 20 in. long. The pitch of the screw is $\frac{1}{4}$ in. What force would be required to lift a car weighing 2000 lb. if the efficiency of the screw is 25%?

Let required force be x lb.wt.

In one complete turn of T-piece, this force acts through distance of $20 \times \frac{22}{7}$ in. $= \frac{20 \times 22}{12 \times 7}$ ft.

$$\therefore \text{work put into machine} = \frac{20 \times 22}{12 \times 7} x \text{ ft.lb.}$$

Now, with one complete turn of T-piece, car rises $\frac{1}{4}$ in. $= \frac{1}{48}$ ft.

Efficiency is 25%, i.e. work got out $= \frac{25}{100}$ of work put in.

$$\therefore \frac{2000}{48} = \frac{25}{100} \times \frac{20 \times 22}{12 \times 7} x$$

Solving this, we get $x = 31.9$ lb.wt.

The Lever. In Chapter 4 of *Junior Physics* we arrived by experiment at the **Law of the Lever**. Where the forces are acting at right angles to the lever we may state it in the simple form:—

$$\text{Load} \times \text{Load Arm} = \text{Effort} \times \text{Effort Arm}.$$

In Fig. 6/7, for instance, we have $w \times \text{FA} = \frac{1}{3}w \times \text{FB}$. To include the kind of case represented in Fig. 6/8, we

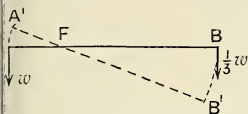


FIG. 6/7

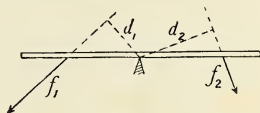


FIG. 6/8

extended the more general statement: *When a lever is at rest under the action of two forces, the product (force) \times (perpendicular distance from the fulcrum to its line of action) is the same for both sides.* Mathematically, the law may be expressed by the equation $f_1 d_1 = f_2 d_2$.

Let us consider a simple lever in action—Fig. 6/7, for instance, can represent a crowbar AB with its fulcrum at F. An effort applied at B overcomes a resistance at A, and the lever moves into the position A'B'. The work put into the machine is the force at B multiplied by the distance BB'. The work got out of it is the *larger* resistance at A multiplied by the *smaller* distance AA'. It is quite an easy exercise in similar triangles to prove that the two amounts of work are equal—i.e. the Principle of Work holds good, as we should expect.

The Balance. A common and important example of the lever is the ordinary beam balance. Here the forces are weights, and we may write $w_1 d_1 = w_2 d_2$. In an accurately made balance $d_1 = d_2$. Hence the d 's cancel and $w_1 = w_2$; that is, the weight of the object is found by adding up the weights in the opposite pan.

If the d 's are not equal, the weight of the object cannot be found in this simple way. In practice we may use what is

called the *method of substitution*. To do this, we put the object in one pan and balance it by means of lead shot, say etc. We then substitute weights for the object, make them balance the lead shot. These weights evidently represent the weight of the object.

Another method is to weigh the object first in one pan and then in the other.

Suppose the left arm of the balance measures d_1 cm., the right arm d_2 cm., and the true weight of the body is W gm. When the body is placed in the right-hand pan it is balanced by w_1 gm. in the left (Fig. 6/10a).

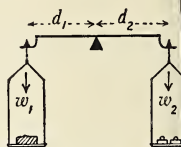


FIG. 6/9

$$\therefore w_1 d_1 = W d_2 \quad . \quad . \quad . \quad . \quad . \quad (i)$$

Again, when placed in the left-hand pan it is balanced by w_2 gm. in the right (Fig. 6/10b).

$$\therefore W d_1 = w_2 d_2 \quad . \quad . \quad . \quad . \quad . \quad (ii)$$

From (i), $\frac{d_1}{d_2} = \frac{W}{w_1}$, and from (ii) $\frac{d_1}{d_2} = \frac{w_2}{W}$.

$$\therefore \frac{W}{w_1} = \frac{w_2}{W}, \text{ giving } W^2 = w_1 w_2, \text{ or } W = \sqrt{w_1 w_2}.$$

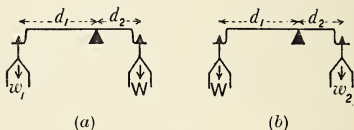


FIG. 6/10

Thus the true weight of the body is the *geometric mean* of the apparent weights on the two different sides. In practice the arithmetic mean, $\frac{1}{2}(w_1 + w_2)$, is taken, as there is appreciable difference between this and the geometric mean provided w_1 and w_2 do not greatly differ.

Problems on balances with unequal arms are easily solved by applying the law of the lever, as in the following.

Example 5. The right arm of a balance is 15.00 cm. long, the left arm 15.06 cm., and the beam balances when no scale-pans are attached. If it also balances when the scale-pans are attached, and one on the left weighs 50 gm., calculate the weight of the other pan. Suppose an object is placed in the left-hand pan and the beam is horizontal when a 100 gm. weight is placed in the other pan, what is the weight of the object?

If R is the weight of the right-hand pan, we have (by the law of the Lever),

$$R \times 15 = 50 \times 15.06,$$

which gives $R = 50.2$ gm.

If x is the weight of the object, the total weight on the left is $50 + x$, and the total weight on the right is $100 + 50.2 = 150.2$.

\therefore by the law of the Lever,

$$(50 + x) \times 15.06 = 150.2 \times 15.00.$$

Solving this equation, we obtain $x = 99.60$ gm.

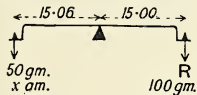


FIG. 6/11

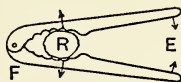


FIG. 6/12

Other 'orders' of levers. In all the levers mentioned in *Senior Physics*, and in the balance just discussed, the fulcrum lies between E and R , where E is the point of application of the effort and R of the resistance. Such levers are often said to be of the first 'order' or class.

In many cases, however (e.g. a pair of nut-crackers), R lies between E and F . Such a lever is said to be of the *second order*.

Example 6. In Fig. 6/12 suppose $RF = 2$ in. and $EF = 5$ in. We apply a force of 6 lb. weight at E , what pressure is experienced the nut at R ?

If the required pressure is x lb. weight we have by the law of Lever,

$$x \times FR = 6 \times EF$$

$$\therefore x \times 2 = 6 \times 5$$

$$\therefore x = 15 \text{ lb.wt.}$$

One is apt to become rather confused in applying the law of the Lever in a case of this sort, and it is perhaps better to think in terms of the Principle of Moments, though it really amounts to the same thing. Thus taking moments about F we have

$$\text{clockwise moment} = 6 \times EF$$

$$\text{anticlockwise moment} = x \times FR.$$

$\therefore x \times FR = 6 \times EF$, and we finish as before.

Levers of the *third order*, i.e. in which E lies between F and R are met with in sugar-tongs and a common pattern of fire tongs. The pedal of a knife-grinder's machine, or of a sewing machine, is also a lever of this kind, and so is the safety-valve of a steam-engine (Fig. 6/14).

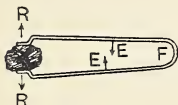


FIG. 6/13

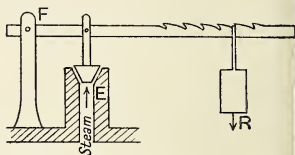


FIG. 6/14

Another interesting example is the human fore-arm. The elbow, may be regarded as the fulcrum, the load rests in the hand at R, and the effort is applied at E, where the lower end of the biceps muscle is attached.

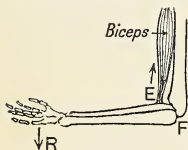


FIG. 6/15

Notice that in levers of the third order, the force ratio is always less than 1, because the effort arm is always shorter than the load arm. Suppose for instance, that at E (Fig. 6/15), we have to apply an effort e to overcome resistance r (practically a 'load') at R. Taking moments about F we have

$$\text{clockwise moment} = e \times EF.$$

$$\text{anticlockwise moment} = r \times FR.$$

$$\therefore \frac{\text{load}}{\text{effort}} = \frac{r}{e} = \frac{EF}{RF}.$$

But EF is less than RF.

$$\therefore \frac{r}{e}, \text{ the force ratio, is less than 1.}$$

Correspondingly of course the velocity ratio is less than 1. Thus the up-and-down distance moved by the knife-grinder's foot for instance is less than that moved by the end of the pulley—an arrangement which obviously makes for comfort. Similarly, a large up-and-down movement of one's fore-arm is secured by a much smaller contraction and elongation of the biceps.

Questions

A

1. Fill in the blanks in the following table:—

Machine	Effort (lb. wt.)	Dis- tance moved by effort (ft.)	Load (lb. wt.)	Dis- tance moved by load (ft.)	Velo- city ratio	Force ratio	Effi- ciency (%)
Wheeler's	...	2	99	1	99
Block and tackle (fig. 6/2)	12	16	...	4	75
Inclined plane (fig. 6/4)	20	5	$6\frac{1}{4}$	5	...

2. Using the block and tackle of Fig. 6/2 on p. 52 a man is able to raise a load with an effort of $12\frac{1}{2}$ lb. wt. and by pulling in 8 ft. of rope to raise a load weighing 45 lb. to a height of 2 ft. Find (i) the velocity ratio, (ii) the work got out of it, (iii) its efficiency.

3. Find the force ratio and the velocity ratio of the machine in Fig. 6/2, and hence find the efficiency. Compare your answer with that previously obtained.

B

1. Describe how you would determine experimentally the efficiency (for one load) of a pulley system consisting of one fixed and one movable block. The number of pulleys in each block is to be assumed unknown.

2. A sack weighing 1 cwt. is lifted 30 ft. by such a system in which the mechanical advantage is four and the efficiency 80%. Find the effort applied and the work done by it in lifting the sack. Neglect the weight of the lower block. Lond.

2. Define the *efficiency*, *mechanical advantage* and *velocity ratio* of a machine.

Give a labelled diagram of a machine suitable for lifting a load of about 100 lb. wt. when a force of about 25 lb. wt. is available to work it, and explain how it would be used. State the approximate value of the velocity-ratio of the machine and describe briefly how you would measure it. *Lond.*

3. A capstan is 1 ft. in diameter and has four bars at right angles to the axle. One man pushes against each bar and exerts a force of 50 lb. wt., so as just to raise an anchor weighing 1000 lb. wt. How far along the bars from the centre of the capstan's axle must each man push? *O.C.**

4. Explain what is meant by the *mechanical advantage*, *velocity ratio* and *efficiency* of a machine, illustrating your answer with reference to the raising of a body by moving it up an inclined plane.

What horizontal force is required to move a barrel of mass 500 lb. up a smooth inclined plane 10 ft. long inclined at 30° to the horizontal, and what is the work done? *Lond.*

5. The following results were obtained in an experiment with a sheaved pulley system with two pulleys in each block:—

Load (lb.wt.)	.	20	40	60	80	100
Effort (lb.wt.)	.	14.3	16.7	22.1	28.6	35.7

Work out the mechanical advantage and efficiency for each value of the load.

Plot a graph of efficiency (Y-axis) against load (X-axis) and comment on the shape of the graph. *Oxf.**

6. Explain the terms *mechanical advantage*, *velocity ratio* and *efficiency* when applied to machines.

Why is 'perpetual motion' considered to be impossible?

In raising a load 1 in. by means of a machine, the point of application of the effort was moved a distance of 90 in. in the direction of the effort. If the efficiency of the machine was 30%, calculate the effort required to raise a load of 600 lb. *Cam.*

7. A screw of pitch 1 in. is worked by a lever 7 ft. in length. If a force of 3 lb. wt. applied at the end of the lever is just sufficient to raise a weight of 1000 lb., calculate (a) the velocity ratio, (b) the efficiency of the machine, (c) the work done by the effort in raising the weight 3 in. (Take π as $\frac{22}{7}$) *O.C.**

CHAPTER 7

COMPOSITION AND RESOLUTION OF FORCES

In *Junior Physics* (Chapter 2) we discussed at some length the meaning of 'Force,' and finally arrived at the definition—*force is that which changes, or tends to change, the state of rest or motion of a body.* In Chapter 6 when dealing with parallel forces we met with the term *resultant*—the **resultant of a number of forces** being *that single force which can take the place of all the separate forces.*

In the present chapter we shall be concerned chiefly with the resultant of non-parallel forces, and we may very well begin by considering the following experiment.

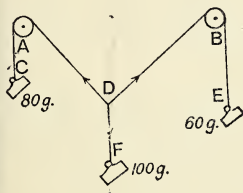


FIG. 7/1

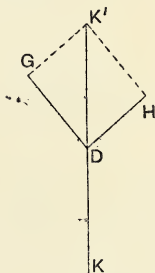


FIG. 7/2

Two light, smoothly-running pulleys A and B are fixed to a board on the wall, and a piece of thread CADBE is passed over them. Weights are attached at the ends C and E, and a third weight F hangs from an intermediate point D by a short branch thread. It does not greatly matter what weights are actually used so long as no one of them is greater than the other two put together, but for the sake of definiteness let us suppose that the weights at C, E and F are 80 gm., 60 gm. and 100 gm. respectively.

At the point D we have three forces acting—viz. (i) 80 gm. along DA, (ii) 60 gm. along DB and (iii) 100 gm. along DF.

If we place a sheet of paper behind the threads we can mark the point that comes just behind D, and we can also mark the directions DA, DB, DF. These directions are shown in Fig. 7/2.

On a scale of, say, 1 in. to 20 gm., represent the force 80 gm., acting along DA. This gives us DG (4 in.). Similarly make DH 3 in. long and DK 5 in. long, to represent forces of 60 gm. and 100 gm. respectively.

Produce KD to K' making $DK' = DK$.

Now, as DK' is equal and opposite to DK, it represents an upward force that would just balance the downward force 100 gm. (i.e. it represents the *equilibrant* of the downward force—see *Jun. Phys.*, p. 55). But DG and DH, taken together, *also* represent the equilibrant of the downward force (because in our experiment, the forces acting along DA and DB were, together, able to balance the force along DF). In short,

- (i) the equilibrant of the downward force is represented by DK';
- (ii) the equilibrant of the downward force is also represented by DG and DH.

It follows that the single force represented by DK' can take the place of the two forces represented by DG and DH. In other words, DK' represents the *resultant* of the two forces represented by DG and DH.

If now we join K'G and K'H, we notice that the figure K'GDH is a parallelogram. Our first thought is that this is just a sort of accident, but on repeating the experiment a number of times, constantly varying the weights, we find that the figure K'GDH is *always* a parallelogram. This gives us a ready means of finding the resultant of two forces acting at a point such as D. Draw two lines from D to represent the forces in magnitude and direction. Complete the parallelogram. The diagonal drawn from D will give the required resultant.

Another experiment leading up to the parallelogram rule is this. Three pieces of string DA, DB and DF each about a foot long are knotted together at D (Fig. 7/3) and a spring balance is attached to each. A large sheet of paper is placed

berneath, and a well-marked dot is placed under D. The string balances are now pulled by three boys, one to each balance, the pulls being regulated so that D remains over the dot. When a 'steady' condition has been reached, each boy notes the reading of his balance, while a fourth boy quickly marks the paper to indicate the directions of the strings.

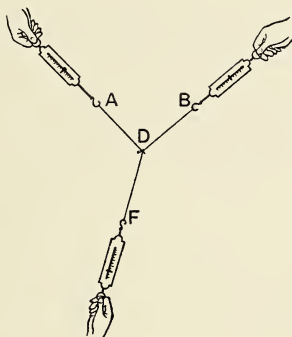


FIG. 7/3

We now have three forces in equilibrium, the magnitude and direction of each being known, and the results are developed in the way already described.

The conclusion which can be reached by experiments such as those we have described is known as the **Parallelogram of Forces**. *If two forces acting at a point are represented in magnitude and direction by adjacent sides of a parallelogram, their resultant will be completely represented by that diagonal of the parallelogram which passes through the same point.*

This statement should be carefully committed to memory.

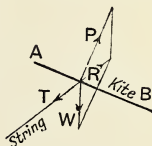


FIG. 7/4

A good example of the parallelogram of forces is seen in the flying of a kite. In Fig. 7/4 AB represents the kite seen sideways on. P represents the effective force of the wind's pressure, and W the weight of the kite. By

the parallelogram of forces we can find R , the resultant of P and W . For equilibrium, T , the tension in the string, must be equal and opposite to R .

Let us now work through one or two numerical examples in which the Parallelogram of Forces occupies the central place.

Example 1. *There is a ship aground, and a tug is pulling it towards the east with a force equal to 25 tons weight, while another is pulling north-east with a force equal to 15 tons weight.*

Suppose a third tug is to take the place of these two, how hard must it pull, and in what direction, so as to produce the same effect?

Choosing a suitable scale (say 1 cm. = 5 tons) draw two lines AB and AC to represent forces of 25 tons to the east and 15 tons to the north-east respectively.

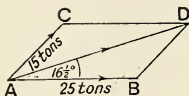


FIG. 7/5

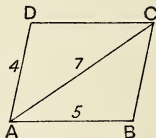


FIG. 7/6

Draw CD parallel to AB and BD parallel to AC . The resultant is represented by AD . By careful measurement we find it to be 7.42 cm. representing $7.42 \times 5 = 37.1$ tons, while $\angle BAD$ is found to be $16\frac{1}{2}^\circ$.

Thus the answer is **37.1 tons weight in a direction $16\frac{1}{2}^\circ$ N of E**

Example 2. *Two forces of 5 lb. and 4 lb. weight respectively act at a certain angle. Find this angle, given that the resultant is a force of 7 lb. wt.*

If we imagine the complete parallelogram of forces to be drawn it would appear somewhat as shown in Fig. 7/6. Clearly the side AD would be 4 units long (because opposite sides of a parallelogram are equal). This suggests the necessary construction—

Make a triangle ABC , with $AB = 5$ units of length, $BC = 4$ units, and $AC = 7$ units. Complete the parallelogram $ABCD$. Then $\angle BAD$ is the required angle, and is found by measurement to be 76° .

Triangle of Forces. Let us go back to Fig. 7/2 (for convenience, repeated as Fig. 7/7) in which DG , DH and DE represent three forces in equilibrium. Now, these three

es are represented in magnitude and direction (but not regards their point of application) by the three sides of triangle DHK'. For comparing the sides of this triangle in DH, DG and DK, we see that—

is unchanged.

' = DG and is in the same direction, though it is not drawn from the point D.

D = DK and is in the same direction, but drawn *towards* the point D instead of away from it.

he situation is summed up by a statement known as the **Triangle of Forces**, viz. *If three forces acting at a point are in equilibrium, they can be represented in magnitude and direction (though not in position) by the three sides of a triangle taken in order.*

Three sides of a triangle taken in order means that the end-point of one side is regarded as the starting point of the next. Thus after DH (ending at H), the next side is HK', not K'H.

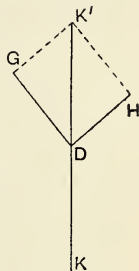


FIG. 7/7

Example 3. *A mass of 6 lb. hangs from the end of a light string, 4 ft. long. What force, applied by another string attached to the same point of the mass and inclined at 45° to the horizontal, will be required to pull the mass aside to a distance of 2 ft. from the vertical through the point of suspension? (The problem may be solved either graphically or by calculation.) J.M.B.**

The mass of 6 lb. is in equilibrium under the action of three forces. (i) Its own weight acting along BC, vertically downwards; (ii) the tension of the 4 ft. string, acting along BA; (iii) the tension of the other string acting along BD (Fig. 7/8a).

Since these forces are in equilibrium, they can be represented by the three sides of a triangle (except as regards their point of application). Accordingly (Fig. 7/8b), draw PQ vertically upwards, and 6 units long (to represent 6 lb.). From Q, draw QR parallel to BD, and from P draw PR parallel to AB. We are asked to find the force along BD. This is represented

Strictly, after PQ and QR we should draw RP (not PR), so that the sides would be 'taken in order.' As P is already fixed in position, it is more convenient to draw PR, but of course the force along BA is represented by RP (not PR).

by QR, which is found to measure 3.11 units. The required force is therefore **3.11 lb.wt.**

If required, we could easily (by measuring RP) find the tension in the string BA.

N.B. It is a little easier to let BC itself (extended to length of 6 units) serve, instead of PQ, as one side of the triangle (Fig. 7/8c). We then draw CH parallel to BD, and the third side HB is in line with BA. The required force of course, represented by CH.

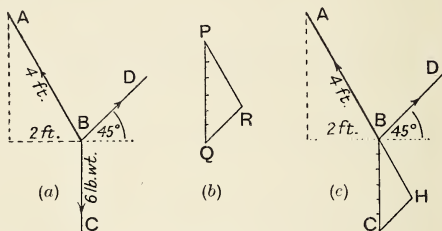


FIG. 7/8

Equilibrium of large body. So far, in this chapter, we have considered the case in which a small body—a point—is in equilibrium under the action of three co-planar forces (three forces in the same plane). We must now consider the case in which the body concerned is of appreciable size. We shall show that in this case the lines of action of the three forces are either (i) all parallel or (ii) all pass through the same point.

In our earlier book we have already had instances in which a body of some size is in equilibrium under the action of *parallel* forces (e.g. the wooden ruler on p. 35). In that case we frequently made use of the fact that the sum of the clockwise moments about any point in the plane must be equal to the sum of the anticlockwise moments, or—more shortly—that the algebraic sum of the moments about any point in the plane must be zero. If this were not so, the body could not be in equilibrium—it would rotate.

Suppose, however, the lines of action of the three forces (say A, B and C in Fig. 7/9) are not all parallel. Then

st two of them, say A and B, must meet if produced. Let them meet at O. The two forces A and B may be replaced by a single resultant, R. Find this resultant by the parallelogram of forces or otherwise.

The body is now in equilibrium under the action of *two* forces, C and R, which must therefore be equal and opposite,¹ they must have a resultant of zero.

But the line of action of R passes through the point at which A and B intersect.

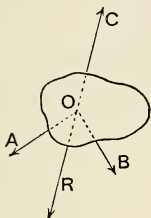


FIG. 7/9

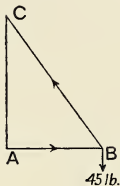


FIG. 7/10

Therefore C (which is in the same straight line as R) also passes through the point at which A and B intersect—i.e. A, B, and C all pass through the same point.

Summing up, if a body is in equilibrium under the action of three non-parallel co-planar forces, the lines of action of these forces must meet in a point, and the resultant of the forces must be zero.

Let us now consider some further examples related to the Equilibrium of Forces.

Example 4. Explain the rule known as the parallelogram of forces, and show how it can be tested experimentally.

A horizontal pole 3 ft. long, with one end hinged to a vertical wall, supports, by a wire from its other end, a large electric sign weighing 45 lb. Fastened to the end of the pole remote from the wall is a

If this is not clear, think of a board with two pegs to which strings are attached. The strings are pulled. Obviously the board will move until the strings are in the same straight line.

Again, unless the forces along the strings are equal, the board will move in the direction of the greater force.

Thus the board will not be in equilibrium unless the two forces are equal and opposite.

supporting wire 5 ft. long, fixed to the wall vertically above the pole. Find graphically the forces acting in the pole and wire. J.M.B.

For first part of question see pp. 65-67.

The arrangement of the pole, etc., is shown in Fig. 7/10, where AB represents the pole and BC the wire.

At B we have three forces in equilibrium

- (i) wt. of electric sign, 45 lb. downwards,
- (ii) tension of wire acting along BC,
- (iii) thrust of pole acting along AB.

Since the forces are in equilibrium, we can represent them the three sides of a triangle taken in order.

It is usually necessary to draw such a triangle, but in Fig. 7 it already exists, viz. the triangle ABC. (CA representing weight of 45 lb., AB the thrust and BC the tension.)

Now, AB = 3 units of length, BC = 5 units and by measurement or calculation CA = 4 units.

4 units of length represent 45 lb.

$\therefore 1 \text{ unit} \text{ represents } 11\frac{1}{4} \text{ lb.}$

$\therefore \text{AB (3 units) represents } 11\frac{1}{4} \times 3 = 33\frac{3}{4} \text{ lb.}$

and BC (5 units) represents $11\frac{1}{4} \times 5 = 56\frac{1}{4} \text{ lb.}$

Thus a thrust of $33\frac{3}{4}$ lb.wt. acts along AB and a tension $56\frac{1}{4}$ lb.wt. along BC.

Resolution of Force. We have seen how a number of forces can be compounded to give a single resultant. Conversely a given force may be 'resolved' into a number of components (in practice usually two), acting in specified directions.

Suppose, for instance, we have a force of 5 lb.wt. acting to the south, and we wish to resolve it into two components, one acting 20° E. of S. and the other 30° W. of S.

Draw OA 5 units long to represent the 5-lb. force acting to the south, and draw OB and OC in the specified directions.

From A draw AP parallel to CO cutting OB in P, and AQ parallel to BO meeting OC in Q. By parallelogram of forces, OA represents the resultant of forces represented by OP and OQ. Conversely, OP and OQ represent the *resolved parts*, in the directions OB and OC, of the force represented

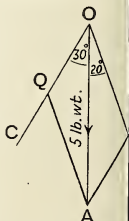


FIG. 7/11

OA. By measurement, OP is found to represent a force of 3 lb.wt., and OQ one of 2.2 lb.wt.

We can illustrate the process experimentally by a very slight modification of the 'spring-balance' experiment described on p. 67.

Draw OA, OB and OC in the specified directions, and produce AO to A' (Fig. 7/12). Three boys now take spring balances attached to knotted strings as already described. The knot is placed over O and the balances are extended along OA', OB and OC respectively. The reading along OA' is adjusted to 5 lb., and the boys pulling in the other



FIG. 7/12

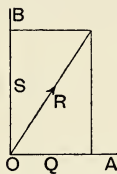


FIG. 7/13

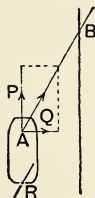


FIG. 7/14

directions increase or reduce their tensions accordingly. It could be found, of course, that the reading along OB is 3 lb., while that along OC is 2.2 lb.

Components at right angles. It is often necessary to resolve a given force into two components at right angles to one another; we shall have examples presently.

Notice that when we speak of the resolved part of a force along a given line, it is understood that the other component is at right angles to that line. Thus in Fig. 7/13 Q is the resolved part of R along OA.

Illustrations. There are numerous cases in daily life in which we have to consider the resolution of a force into two components at right angles. For instance, you have seen a barge towed along a canal from the tow-path, and it perhaps seems curious that it preserves a movement parallel to the bank instead of following the direction of the tow-rope.

But suppose AB is the tow-rope. The tension along A may be resolved into

- (i) a force P parallel to the bank, which pulls the boat forward,
- (ii) a force Q at right angles to the bank. This gives a tendency to turn inwards, which is counteracted by a suitable setting of the rudder (Fig. 7/14).

Somewhat similar is the case of a boat sailing into the wind. Suppose AB is the boat's sail and the wind is blowing from the direction CD. Resolve the wind's force into P acting at right angles to the surface of AB, and Q acting along it. As there is not much friction the effect of Q may be neglected. Now resolve P into two components, S at right angles to the length of the boat and T in a forward

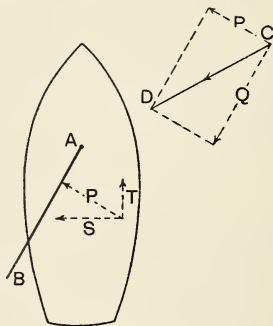


FIG. 7/15

direction. It is T that drives the boat along. S causes the boat to 'make leeway,' but this is reduced by the use of a keel 'centre-board.'

The aeroplane. Suppose an aeroplane is taking off, and owing to the action of the propellor is moving rapidly from left to right along the ground. Relative to the plane, the air of course moves from right to left. Chiefly owing to the shape of the wing, there is a region of high pressure formed below it and of low pressure above it, and the net effect is to produce an upward force (represented by F in Fig. 7/16).

ly at right angles to the wing. We can resolve F into vertical force L and a horizontal one D . L is known as *lift*, and if the plane is to rise it must obviously be a little greater than the total weight. D is the *drag*, and resists the forward movement of the aeroplane. It has to be overcome by the pull of the propellor.

The magnitude of L evidently depends on the magnitude of F , which in turn depends on how fast the aeroplane is being moved. If the movement is too slow L will not be great enough to lift the aeroplane. We have the same principles at work when a boy is beginning to fly his kite. In this case, he runs along the ground instead of setting the propellor to work.

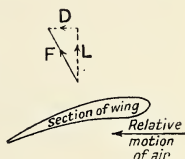


FIG. 7/16

Questions

A

By a graphical method or otherwise, find the resultant of the following forces:

- (i) 3 lb.wt. and 4 lb.wt. acting at an angle of 90° .
- (ii) 20 gm. „ 30 gm. „ „ „ 180° .
- (iii) 20 gm. „ 30 gm. „ „ „ 0° .
- (iv) 3 lb. „ 3 lb. „ „ „ 60° .
- (v) 10 lb. „ 10 lb. „ „ „ 45° .

Two forces, one of them of 6 lb.wt., act at right angles and give a resultant of 10 lb.wt. By a graphical construction, or otherwise, find the magnitude and direction of the other force.

A common method of hanging a picture is by means of a cord passing like an inverted V over a nail. Chiefly by means of a force diagram, point out the forces that are in equilibrium at the point of suspension.

B

State the theorem of the parallelogram of forces and describe an experiment to verify it.

The resultant of two forces is of magnitude 150 lb.wt. and makes angles of 25° and 60° with them. Find, by a graphical method, the magnitudes of the forces.

An unknown mass is suspended by a light cord and is acted on by a horizontal force of 100 gm.wt. In the position of

equilibrium the cord makes an angle of 30° with the vertical. Draw a force diagram and find the value of the mass. *Lond.*

3. Under what conditions will (a) two forces, (b) three forces acting at a point, be in equilibrium?

Three spring balances are attached to a small ring. Two are pulled horizontally, one due north and the other south-east, the respective readings being 8 lb.wt. and 12 lb.wt. In what direction must the third spring balance be pulled in order that the ring may be kept at rest? What will be the reading on this balance? *Lond.*

4. The bob of a simple pendulum, hanging vertically, has a mass of 30 gm. A silk thread is attached to the bob and pulled horizontally. If the silk breaks when the tension in it exceeds 10 gm.wt., find (a) the maximum inclination of the pendulum with the vertical, (b) the tension in the string of the pendulum in that position. *Lond.**

5. Why is it better to tow a barge along a canal, from the tow-path, by using a long rope rather than a short one? *Lond.*

6. To the ends of a straight uniform beam of length 4 ft. and mass 5 lb. are secured the ends of a cord 8 ft. long. The beam is then hung in a horizontal position by passing the cord over a fixed hook. What is the tension in the cord? *Dur.**

7. Explain what is meant by the *triangle of forces*. Give a brief description of an experiment, carried out by you, to verify the principle.

A uniform bar of metal AB, weighing 80 lb., is suspended by two strings, AC and BC, fastened to a nail at C. If the strings make angles of 45° and 30° respectively with the horizontal, determine graphically, or otherwise, the tension in each string. *C.W.B.*

8. A mass of 50 lb. is supported by two threads inclined respectively at angles of 60° and 30° to the vertical. Calculate the tensions in the threads. *O.C.**

9. Forces of 50, 40 and 30 gm.wt. act at a point in directions north, east and south respectively. Find their resultant magnitude and direction. *Lond.**

10. A uniform rod, AB, 5 ft. long and weighing 10 lb., is suspended by strings OA and OB. $OA = 4$ ft., $OB = 3$ ft. On squared paper, using a scale of 1 in. = 2 lb., draw a diagram of the forces acting on the rod when it is at rest, and mark clearly what each line represents. By direct measurements, find the tension in each string. *Camb.**

11. A kilogram weight is placed on a smooth plane inclined to the horizontal at 60° and is kept from sliding down by a string parallel to the plane. Calculate the tension in the string and the reaction between the weight and the plane. *Camb.**

CHAPTER 8

MOMENT OF A FORCE. CENTRE OF GRAVITY

This chapter will consist largely of a revision of the material treated in Chapters 5 and 6 of *Junior Physics*, but we shall consider some examples that would have been too difficult for inclusion in the earlier book.

First, then, let us recall the meaning of 'moment of a force.' **The moment of a force about a fixed point** is the product of the force itself, and the perpendicular distance from the fixed point to the line of action of the force.

Thus in Fig. 8/1, P is the fixed point, and the force F is

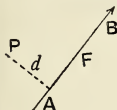


FIG. 8/1

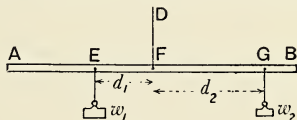


FIG. 8/2

represented by the straight line AB. If the perpendicular distance from P to AB is d , then the moment of F about the point P is given by the product $F \times d$.

The moment of a force about a point as just defined is a measure of the *turning powers* of the force about that point. This is best seen by considering a lever, balanced by forces represented by weights. In the case indicated in Fig. 8/2 it is shown by experiment that $w_1 \times d_1 = w_2 \times d_2$ —i.e. the *moments* of the forces represented by w_1 and w_2 are equal.

But the *turning powers* of these two forces are also equal, w_2 tends to turn the lever clockwise and w_1 tends to turn it anticlockwise; and as the lever remains horizontal, these tendencies must be equal and opposite. Hence the moment of a force must be a measure of its turning power.

Principle of Moments. By hanging extra weights on the lever of Fig. 8/3 we could evidently arrange for it to be in equilibrium under the action of more than two forces. Further, though all these 'weight' forces are downwards, we could easily arrange to apply some upward forces by means of

spring balances—such a force is shown as w_5 . If we note the weights (or balance readings) and distances, it would be found that $(w_2 \times d_2) + (w_4 \times d_4) = (w_1 \times d_1) + (w_3 \times d_3) + (w_5 \times d_5)$.

i.e. sum of the clockwise moments = sum of anticlockwise moments.

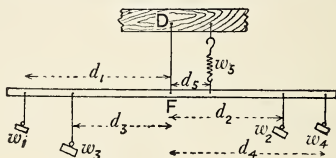


FIG. 8/3

This is an example of the **Principle of Moments** which states that: *When a bar free to move about a fixed point is in equilibrium, the sum of the clockwise moments is equal to the sum of the anticlockwise moments.*

Some examples involving the principles just discussed were given on pp. 30–36 of our earlier book. Here are one or two harder ones.

Example 1. Explain the term moment of a force about a point.

Two uniform rods AB, BC, of different materials, each 50 cm long, are joined at B in a straight line. On hanging a 50-gm weight from A, it is found that the combination balances on a knife-edge 20 cm. from A. When the 50-gm. weight is hung from C, the point of balance is 40 cm. from C. Find the weight of AB and BC.

J.M.B.

For first part of question see p. 77.

Suppose the required weights are w_1 and w_2 respectively. The rods are uniform, and so we may regard w_1 as acting at the middle point of AB, and w_2 at E, the middle point of BC. Let F be the knife-edge.

$$AD = \frac{1}{2} AB = 25 \text{ cm.}; AF = 20 \text{ cm. (given).}$$

$$\therefore FD = AD - AF = 5 \text{ cm.}; FE = FD + DE = 5 + 50 = 55 \text{ cm.}$$

Taking moments about F, we have

$$\begin{aligned} \text{Sum of clockwise moments} &= (w_1 \times FD) + (w_2 \times FE) \\ &= 5 w_1 + 55 w_2 \end{aligned}$$

$$\begin{aligned} \text{Anticlockwise moment} &= 50 \times AF \\ &= 50 \times 20 = 1000 \end{aligned}$$

$$\begin{aligned} \therefore 5 w_1 + 55 w_2 &= 1000 \\ \text{or } w_1 + 11 w_2 &= 200 \end{aligned} \quad \dots \dots (i)$$

Turning now to Fig. 8/4b, it is easily shown that $FE = 15$ cm., $= 35$ cm., while $FC = 40$ cm. (given).

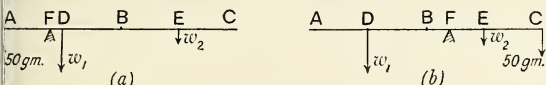


FIG. 8/4

Sum of clockwise moments $= (w_2 \times FE) + (50 \times FC)$

$$= 15 w_2 + 2000$$

Anti-clockwise moment $= w_1 \times DF = 35 w_1$

$$\therefore 35 w_1 = 15 w_2 + 2000$$

$$\text{or } 7 w_1 - 3 w_2 = 400 \quad \dots \quad (ii)$$

Solving (i) and (ii) as simultaneous equations, we obtain $w_1 = 62\frac{1}{2}$ gm. and $w_2 = 12\frac{1}{2}$ gm.

Example 2. A uniform rod of length 8 ft. weight 8 lb. is hinged smoothly at one end to a vertical wall; it is kept in a horizontal position by a string attached to the free end of the rod, to a point in the wall vertically above the hinge and 6 ft. from it. What is the tension in the string? J.M.B.*

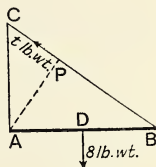


FIG. 8/5

Let AB be the rod hinged at A , AC being the vertical wall.

$AB = 8$ ft. $AC = 6$ ft. $\therefore BC = \sqrt{AB^2 + AC^2} = \sqrt{64 + 36} = 10$ ft.

The rod is in equilibrium under the action of the clockwise moment $8 \times AD$, and the anticlockwise moment $t \times AP$, where t (lb.wt.) is the required tension, and AP is the length of the perpendicular from A to BC .

$$\text{i.e. } 8 \times AD = t \times AP \quad \dots \quad (i)$$

$$\text{Now by similar } \Delta's, \frac{AP}{AC} = \frac{AB}{BC}$$

$$\therefore AP \div 6 = 8 \div 10.$$

$$\therefore AP = 4.8 \text{ ft.}$$

$$\text{Also } AD = \frac{1}{2} \text{ of } 8 \text{ ft.} = 4 \text{ ft.}$$

Substituting these values in (i), we have $8 \times 4 = t \times 4.8$

$$\therefore t = 6\frac{2}{3} \text{ lb.wt.}$$

The problem may also be solved as an exercise in the Triangle of Forces (cf. Ex. 3 on p. 71).

Example 3. A uniform ladder 20 ft. long and weighing 100 stands against a smooth wall with which it makes an angle of 30° . A horizontal force which increases as required up to 60 lb.wt. is applied at the base, towards the wall. How far can a boy weighing 80 go up the ladder before it will slip?

The arrangement is shown in Fig. 8/6.

As the wall is smooth, the reaction AR is at right angles to it. This reaction is equal and opposite to the force at B. The boy is supposed to be at P.

Imagine an axis at B, perpendicular to the paper, and let us take moments about it.

The reaction at A has an anticlockwise moment ($= 60 \times AD$). The weight of the ladder has a clockwise moment ($= 100 \times BH$), and so has the weight of the boy ($= 80 \times BK$). As he moves higher up the ladder, BK increases, and if the clockwise moment should become too great, A will descend and B move outwards, the ladder will slip.

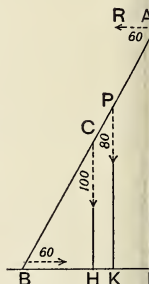


FIG. 8/6

For equilibrium, we must have

$$60 \times AD = (100 \times BH) + (80 \times BK) \quad \dots (i)$$

Now $AD = AB \sin 30^\circ = 20 \times \frac{1}{2} = 10$.

$$BH = \frac{1}{2}BD = \frac{1}{2}AB \cos 30^\circ = \frac{1}{2} \times 20 \times \frac{\sqrt{3}}{2} = 5\sqrt{3}.$$

$BK = BP \cos 30^\circ = x \times \frac{1}{2}$, where x is the required distance up the ladder.

Thus equation (i) becomes

$$60 \times 10 = (100 \times 5\sqrt{3}) + (80 \times \frac{1}{2}x).$$

Solving this, we get $x = 13.5$ ft.

N.B. The force at B is supplied by friction. We shall see later (p. 87) that, up to a certain maximum, this force is always just enough to resist slipping. If it were $= 60$ lb.wt., when the boy began to ascend the ladder, the latter would slip up the wall.

In practice, a man stands at the base of the ladder so that he can add to the horizontal force if necessary.

Let us now recall, and to some degree extend, the work on Parallel Forces and Centre of Gravity which formed the subject of Chapter 6 in *Junior Physics*.

We found that, in general,¹ a group of parallel forces has a single resultant, of which

¹ The exception is the case in which the parallel forces form a couple (p. 82).

- i) the *magnitude* is equal to the sum of the parallel forces,
- ii) the *direction* is that of the parallel forces and
- iii) the *point of application* is such that the sum of the clockwise moments about that point is equal to the sum of the anticlockwise moments.

A simple example is shown in Fig. 8/7, where parallel forces of 4 lb.wt. and 2 lb.wt. are acting at right angles to a bar AB 3 ft. long. The resultant R (i) has a magnitude $4 + 2 = 6$ lb.wt.; (ii) has a direction parallel to that of the given forces and acts at a point C, 1 ft. from A. To find the point of application, use at this point the moment of the 4 lb. force about C (4×1) is equal to the moment of the 2 lb. force (2×2).

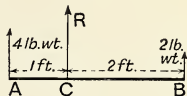


FIG. 8/7

For simplicity the forces in Fig. 8/7 are shown acting at right angles to the bar, but the statements made with regard to the resultant are true whatever the angle between the forces and the bar.

Parallel forces acting in the same sense (e.g. all to the north, all to the east) are said to be *like*. If they act in opposite senses (e.g. some to the north and some to the south) they are said to be *unlike*. The statements made with regard to the resultant still hold good, except that forces acting in opposite directions must be regarded as being of opposite *sign*.

Example 4. Unlike parallel forces of 4 lb.wt. and 2 lb.wt. act at right angles to a bar AB 6 feet long, their points of application are the ends A and B respectively. What will be the direction, magnitude and point of application of their resultant?

Direction. Parallel to the lines of action of the given forces, in the same sense as the greater one.

Magnitude. Calling the given forces $+4$ lb.wt. and -2 lb.wt. magnitude $= +4 - 2 = +2$ lb.wt.

Point of application. Suppose this is at C (Fig. 8/8), where AC = x ft. (C cannot lie between A and B, because in that case the moments would be clockwise). Taking moments about C,

clockwise moments = anticlockwise moments

$$2 \times BC = 4 \times AC$$

$$\therefore 2(3 + x) = 4x$$

$$\therefore x = 3.$$

$$AC = 3 \text{ feet.}$$

Couples. A pair of *equal* unlike parallel forces is known as a **couple**. In this case the forces simply tend to rotate the body to which they are applied, and they cannot be replaced by a single resultant.

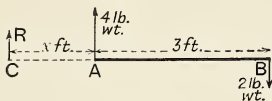


FIG. 8/8

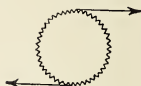


FIG. 8/9

We have a common example of a couple when winding a watch or clock with a key, or a watch with a stem-wind (Fig. 8/9). Another case occurs when we spin a top holding it between the thumb and fore-finger and giving it a sharp turn. Notice that in all such cases it is impossible to secure the same result by applying a *single* force.

Centre of Gravity. In *Junior Physics*, p. 57, we regarded a body as being made up of a great number of separate particles.

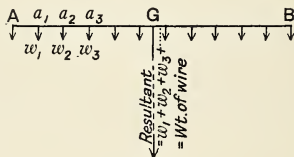


FIG. 8/10

each with its own weight ($w_1, w_2, w_3 \dots$). Thus the body is acted upon by a great number of small, like, parallel forces. The resultant of these is a single force whose magnitude is $w_1 + w_2 + w_3 \dots = W$, where W is the weight of the body and whose direction is the same as that of the parallel forces, i.e. vertically downwards. It was explained further that the whole weight of the body might be regarded as acting through a particular point (G in Fig. 8/10) known as the *centre of gravity*. If we move the body to another place, this point of course, moves with it, but it does not move *relative to the body*. We may now formally define the **centre of gravity** of a body as *that point, fixed relative to the body, through which its weight may always be supposed to act*.

Note the 'always.' In Fig. 8/11a the weight of the body acts through the point H as well as G. If the body be turned round (Fig. 8/11b), the weight still acts through G, but not through H. G is the only point through which the weight always acts.

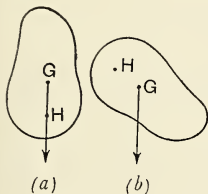


FIG. 8/11

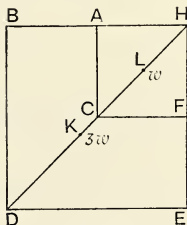


FIG. 8/12

In Chapter 6 of *Junior Physics* we considered certain properties of the centre of gravity, e.g. how its position may decide whether the equilibrium of a body shall be stable, unstable or neutral. We also discussed some practical methods for finding the position of the c.g. We shall now consider an example in which the position of the c.g. is found by calculation.

Example 5. A square of metal, whose sides are 4 in. in length, has one quarter cut away so as to leave an L-shaped piece. How by means of a full-sized diagram the position of the centre of gravity of the piece of metal and explain your construction. How would you check your result by experiment? Lond.

Suppose BDEH was the original square from which the quarter BCHA has been cut away (Fig. 8/12).

If the portion removed weighs w , the remaining L-shaped piece evidently weighs $3w$.

From the symmetry of the figure, we can see that both K the c.g. of the L-shaped piece and L the c.g. of the square ACFH lie on the diagonal DCH.

Imagine that the square ACFH is replaced and that the complete square BDEH, in a horizontal plane, is balanced on a pivoted support placed under C. KCL now becomes our 'weight-lever' balancing at C, because the clockwise moment of the weight w , acting at L, is equal to the anticlockwise moment of the weight $3w$, acting at K.

$$w \times LC = 3w \times KC$$

$$\therefore KC = \frac{1}{3} LC.$$

G

Thus it only remains to measure LC on a full-sized diagram and make CK equal to one-third of it.

The result could easily be checked experimentally, e.g. showing that the sheet of metal balances on the end of a pen placed just below K (this and other methods are fully described in *Junior Physics*, pp. 57-9).

Questions

A

1. A 20-in. scale is suspended from a point half-way along and a 50-gm. weight is hung at a distance of 3 in. from this point. Where would (i) a 20-gm. weight, (ii) a 100-gm. weight have to be placed, to balance it?

2. The 50-gm. weight, placed as above, is balanced by means of a pocket-knife hung at a distance of 4 in. from the fulcrum. Find the weight of the pocket-knife.

3. The scale of q. 1 is suspended at the division marked 1 (in.), but hangs horizontally when a 10-gm. weight is hung at 4 in. Find the weight of the scale.

4. The scale of q. 1 is suspended from its middle point, and weights of 20 gm., 10 gm. and 100 gm. are hung at the points marked 2 in., 4 in. and 8 in. respectively. Where must a 50-gm. weight be hung in order that the scale may balance horizontally?

5. On a sheet of cardboard draw a triangle of sides 4 in., 5 in. and 6 in., and by a geometrical method find its c.g. Now cut out the triangle out, and by balancing it on a pencil find whether you have obtained the point correctly.

B

1. A uniform ladder of mass 200 lb. rests on a pavement against a smooth wall at an angle of 60° to the horizontal. A boy of mass 100 lb. climbs up the ladder, which is prevented from slipping by a force of 100 lb. applied horizontally at the foot of the ladder. By what fraction of the ladder's length can the boy ascend before the ladder slips? *O.C.*

2. Four forces act along the edges of a wooden square ABCD of 6-in. side. The forces are 10 lb.wt. along AB, 15 lb.wt. along BC, 10 lb.wt. along CD and 15 lb.wt. along DA. Find the moment of the couple required to keep the body from rotating. *Oxf.**

3. Explain the meanings of (a) the moment of a force about an axis, (b) the centre of gravity of a body.

A metal box has for its upper side a square lid of uniform thickness hinged at one of its edges. The mass of the lid is 5 lb. and the length of an edge is 4 ft. The lid is opened by applying a vertical force at the edge opposite the hinge. Find (a) the

orce required to incline the lid at 60° to the horizontal, (b) the force at the hinge in this position, (c) the work done in raising the lid from the horizontal to the vertical position. *Dur.*

4. A telescope consists of four concentric tubes, each 10 in. in length, their masses being 4, 3, 2 and 1 lb. Find, by calculation, the position of the centre of gravity. *O.C.**

5. If masses of 1, 2, 3 and 4 lb. are placed at the corners of a square of 10-in. side, taken in order, calculate the position of the centre of gravity. *O.C.**

6. From a thin square plate ABCD of uniform density, and with centre E and sides of 8 in., the triangle AED is cut out. Determine the position of the centre of gravity of the figure which is thus obtained. *O.C.**

7. What are the conditions which must be satisfied when three forces, acting on a small body, keep it in equilibrium?

A uniform rod AB, 10 ft. long and weighing 12 lb., is hinged to a vertical wall at A, and is held at an angle of 30° to the vertical by a horizontal string which is attached to the rod at a point 2 ft. from B and pulls the rod away from the wall. Find graphically (or by calculation), the magnitude and the direction of the reaction at A and the magnitude of the tension in the string. *Camb.*

8. Explain the meaning of the term *centre of gravity* of a body. A body with a flat base is tilted about one edge. Explain the effect of the position of the centre of gravity of the body on the maximum possible angle of tilt before unstable equilibrium is reached.

A compound cylindrical rod 12 in. long is made of two uniform cylinders, each 6 in. long and placed end to end. One is of wood of specific gravity 0.4 and the other of cork of specific gravity 0.24, both being of the same area of cross-section. Calculate the position of the centre of gravity of the rod and say briefly how you would check the position by experiment. *Lond.*

9. Describe how the centre of gravity of an irregularly shaped cardboard sheet can be found experimentally.

A rough aeroplane model is T-shaped, made of cardboard and lies in one plane. The wing span is 50 cm. and its breadth 4 cm.; the length of the fuselage is 40 cm. and its breadth 3 cm.; the weight of the cardboard is 0.20 gm. per sq. cm.; and the weight of the engine, 18.00 gm., may be supposed concentrated at the middle of the leading edge of the wing. The model rests in a horizontal plane when suspended by a single vertical wire. Find the point of attachment of the wire and the tension in it. *J.M.B.*

CHAPTER 9

FRICTION

FRICTION is the force called into play when one surface moves over another. Thus we experience friction when we drag a desk across the school floor, or when we try to pull a nail out of a piece of wood. In our previous work we have often found it necessary to refer to this force, as in Chapter 4 where we considered the part that it played in the act of walking, in the case of a horse pulling a cart, etc. We must now study the subject more closely.

If friction ceased. It is instructive to reflect for a little while on what would happen if the force of friction suddenly ceased to exist. Suppose you were on a level road on your way to school when the change occurred. You would no longer be able to walk. You might fall, or you might remain on your feet, but in either case you would continue to move forward in the same straight line (Newton's 1st law) until you met with some obstacle—a wall, or some unfortunate moving in the opposite direction. If you happened to be going downhill at the moment when friction ceased, you would move with ever-increasing speed until you met with an obstacle. If going uphill, however, you would soon begin moving backwards—you would not be able to turn round. Perhaps the most happily-placed boy would be one who had happened to turn round and stop just before the sudden collapse of friction. By blowing gently he would be able to move in the required direction.

Things would be little better if friction had ceased in the course of a lesson. The master in the act of writing on the blackboard would move backwards towards his class, but the boys would not be in a position to enjoy such a delicious situation because the desks on which they were sitting would have collapsed—for nails and screws owe their grip to friction. Not many of the fallen would wish to pursue their studies amid the wreckage, but if they did, the fact that lead pencils had slid out of their wooden sheaths and pen-nibs from their holders would add to their troubles. However, you can

continue the theme for yourself. Let us get back to the world we know.

Friction and heat. In the case of a machine, we have seen again and again that work put in is (at least to a large extent) presented by work got out. But when we do work in overcoming friction there is, so to speak, nothing to show for it. Actually it is converted into heat.

To raise the efficiency of a machine, we must evidently arrange that as little work as possible shall be spent in overcoming friction, and hence the need for *lubrication*. Instead of allowing one metal surface to rub against another, we arrange that the two surfaces shall be separated by a thin film of oil or grease.

Limiting friction. Suppose a wooden block is resting on a table, with a spring balance attached to it. We pull with a

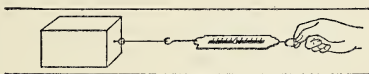


FIG. 9/1

force, say, of 20 gm., and the block does not move. Why not? Evidently because a frictional force of 20 gm., in the opposite direction, has been called into play.

We increase the forward pull to 25, 30, 35 gm., and still the block does not move. Friction is evidently a force which can vary, within limits, so as to be exactly equal and opposite to the forward pull on the body concerned.

But when this forward pull has been increased to a certain amount—say, 40 gm.—the wooden block begins to move. The friction of which we have been speaking has a limiting value, and this is known as the *limiting friction*, or *limiting static friction*.

Once the block has begun to move forward, we shall probably notice a sudden slight slackening in the pull of the spring balance, perhaps to 36 gm. Thus, though a pull of 36 gm. is not sufficient to set the block in motion, once it is in motion this pull is sufficient to keep it moving. The force required to keep a body moving is a measure of what is called

the *kinetic* or *sliding* friction, and from what has been said we see that it is somewhat less than limiting statical friction.

Explorers in polar regions may find it very difficult to get a sledge to move, but once it has started it is not so hard to keep it moving. In the first case they have to exert a force great enough to overcome statical friction; in the second only the smaller force required to overcome sliding friction is needed.

Laws of friction. From common experience we know that the force required to drag a load along a level floor has to be increased if the load is increased. Is the force *proportional* to the load?

We can answer this question by the help of a spring balance but better results can be obtained by using a pulley and balance pan as follows.

A wooden block is attached to a scale-pan by a thread passing over a pulley¹ as shown in Fig. 9/2. The weight of the scale-pan (nearest gram) is first found. Weights are then placed in it until the block just begins

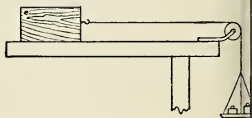


FIG. 9/2

to move. Increasing loads are placed on the block, and the minimum force required to produce motion is observed in each case. The results may be set out as follows:—

Wt. of block and added load (R).	Force required to move block etc. (F).	$\frac{F}{R}$

R stands for 'normal reaction,' i.e. the reaction between table and block in a direction at right angles to the surface in contact. In this case it is equal to the total weight.

¹ In practice it will be found convenient to use the inclined plane (*Junior Physics*, p. 38) lying horizontally, instead of the actual table-top, the pulley of course remaining attached to the plane.

It will be noticed that F/R is approximately constant, i.e. *the limiting friction is proportional to the normal reaction*.

Now turn the block on to one of its narrow faces, and repeat the experiment, using the same loads as before. We find that we obtain the same values for F , i.e. *the friction is independent of the area of the surfaces in contact*.

We have a practical illustration of both these laws in the case of the rim brake of a bicycle. The braking effect is, of course, due to the friction between the brake-shoes and the rim. By gripping the brake-lever more tightly we increase the normal reaction,¹ and the friction in proportion.

We do *not* increase the braking effect, however, by increasing the size of the brake-shoes—though, by way of compensation, large shoes would not wear out so quickly as small ones.

Coefficient of friction. The experiment of p. 88 showed that for a given pair of substances the ratio F/R is constant. This ratio, often denoted by the Greek letter μ (mu), is called the **coefficient of friction**.² Its value for wood on wood (dry) is between 0.25 and 0.50, while if the wood is soaped it is about 0.10. For metal on metal (dry) it is between 0.15 and 0.20, but this may be reduced by oiling to 0.05 or less—a point of interest to cyclists.

Rolling friction. A cart in hilly country is sometimes fitted with a 'skid brake.'

The wheel is now locked. It slides instead of rolling, and the braking effect is very marked. The driver is making use of the fact that sliding friction is greater than rolling friction. For iron wheels rolling on iron rails the coefficient of friction may be as little as 1/100th part of the coefficient of sliding friction.

You may wonder why there should be any friction at all in such a case. The fact is that however firmly the rail may

¹ When considering the experiment of the wooden block and scale-pan, the reader may have thought that 'weight' would be a simpler term than 'normal reaction.' The latter term is necessary, however, as to cover cases in which (as in the bicycle brake) friction is not due simply to weight.

² Strictly speaking, coefficient of *statical* friction, because in our experiment we found the force F necessary to start the block into motion (from rest). The force necessary to keep it moving, when once started, would be somewhat less, as already mentioned. Calling this F' , F'/R would be the coefficient of *sliding* friction.

be laid, it always 'gives' a little, as the wheel passes over it (Fig. 9/3, where the effect is purposely exaggerated). Thus the wheel has to ascend the little hill AB, and in fact an unending series of such little hills.

In machinery of all sorts it commonly happens that one surface is revolving in contact with another. A simple example is the socket of a cart wheel revolving about the end of the axle. Here we have sliding friction. We may arrange, however—in the bicycle wheel for instance—to have a layer of steel balls between the two surfaces, as in the illustration on the left. The balls roll round between the two surfaces, and we have rolling friction instead of sliding friction, the device being known as 'ball bearings.' One often sees horse-drawn milk-carts, etc., in which the wheels and axle (sliding friction) have been re-

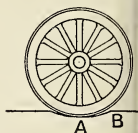
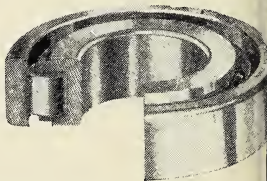
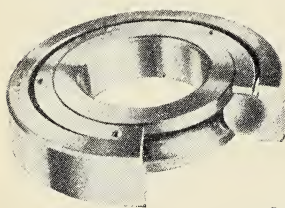


FIG. 9/3



By courtesy of the Hoffmann Manufacturing Co., Ltd.

BALL BEARINGS AND ROLLER BEARINGS

moved and replaced with the wheels of an old motor-car (rolling friction). Such a vehicle is not only much quieter but can be pulled with a great deal less effort.

For heavier machinery, such as a motor lorry, roller bearing may be used, instead of ball bearings. Some of the most up-to-date trains have been fitted in this way.

Questions

1. What do you mean by *friction*? Give three everyday examples.

2. Supposing friction ceased to act, what differences would you notice in carrying out the following operations: (a) walking; (b) nailing up a packing-box; (c) digging a sand-pit by the sea-shore; (d) rubbing your hands together on a cold morning to make them warm; (e) striking a match?

3. Explain the meaning of *coefficient of friction*.

In the experiment described on p. 88 the rectangular wooden block weighed 360 gm. and the scale-pan 20 gm. When a 100 gm. weight was placed in the scale-pan the block was just able to move uniformly. What value does this give for the coefficient of sliding friction?

If weights totalling (i) 90 gm., (ii) 150 gm. were placed on the block, what weights would have to be placed on the empty scale-pan to maintain uniform motion?

4. Explain what you understand by *friction*, and describe briefly the experiments you would perform to illustrate the laws of sliding friction.

Why is it more dangerous to apply the brakes of a car suddenly when it is on a slippery surface than when it is on a rough surface?

Camb.

CHAPTER 10

PRESSURE IN LIQUIDS AND GASES

A. Liquids

THERE have been many cases in which a vessel carrying valuable cargo, perhaps including gold bullion, has sunk in deep water, say 1000 ft. or more. As a rule the cargo in such a case is a total loss, while if the vessel had sunk in shallow water it would soon have been recovered by divers. The point is, of course, that a diver can stand only a limited amount of pressure—and pressure increases with depth. Our earlier work has taught us that it also increases with the density of the liquid, the exact amount (in gm. per sq. cm.) being given by the product hs , where h = depth below surface (in cm) and s = specific gravity of liquid. Usually, however, there is a pressure *at* the surface—that of the atmosphere—and this must be added. If π gm. per sq. cm. is the pressure at the surface, the pressure at depth h becomes $\pi + hs$ (Fig. 10/1).

In the case of water, $s = 1$, and the pressures just given become h and $\pi + h$, respectively.

Further, we learnt that the pressure acts equally in all directions, and that the *resultant* pressure is at right angles to any surface in contact with the liquid. Thus, when water is squirting out of a leaky hose-pipe, each escaping jet is at right angles to the surface of the tube at that point.

We saw that increase of pressure with depth has many practical consequences. It is closely connected with the distribution of water from a town reservoir, and with the formation of artesian wells. It also enters into a multitude of subjects, such as the diving bell, the diver's dress and the submarine.

Here is a numerical problem related to the subject under discussion.

Example 1. A Spanish diver off Cape Finisterre has worked at a depth of 182 ft. Calculate the pressure in lb. per sq. in. (Atmospheric pressure = 15 lb. per sq. in.; 1 cu. ft. of sea-water weighs 65 lb.)

The chief difficulty here is that of *units*, sq. ft. instead of sq. in. and so on.

It will be best to think of a cylinder of sea-water 1 sq. ft. in cross-section placed as shown in Fig. 10/2, i.e. its upper end is at the surface and its lower end 182 ft. below.

Volume of cylinder of water = 182 cu. ft.

weight = 182×65 lb.

This weight is pressing on 1 sq. ft.

\therefore pressure in lb. per sq. in. = $\frac{182 \times 65}{144} = 82.1$.

To this must be added the pressure of the atmosphere, giving a total pressure of $82.1 + 15 = 97.1$ lb. per sq. in.

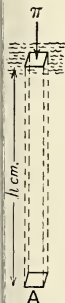


FIG. 10/1

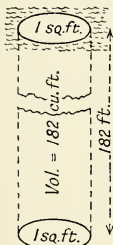


FIG. 10/2

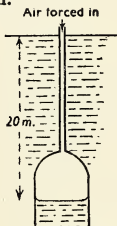


FIG. 10/3

Example 2. The free surface of water in a diving bell is 20 m. below the surface of the sea. The atmospheric pressure is 75 cm. of mercury. What will be the height of a barometer in the diving bell? (Sp. gr. of mercury is 13.6 and of sea-water is 1.025.) Camb.*

f. Fig. 10/3. 20 m. = 2000 cm.

Pressure at this depth (omitting atmospheric pressure) = hs gm. per sq. cm. = $2000 \times 1.025 = 2050$ gm. per sq. cm.

We must add the atmospheric pressure in gm. per sq. cm.

Now, the height of a barometer is the same whatever the bore of the tube, so for convenience suppose the bore to be 1 sq. cm.

Then barometer contains $75 \times 1 = 75$ c.c. of mercury.

Wt. of this = $75 \times 13.6 = 1020$ gm., and all this is pressing down on 1 sq. cm.

Hence barometric pressure = 1020 gm. per sq. cm.

\therefore total pressure in diving bell = $2050 + 1020 = 3070$ gm. per sq. cm.

Our problem is now:—If a barometer is 75 cm. high when the pressure is 1020 gm. per sq. cm., what is its height when the pressure is 3070 gm.?

$$\text{Required height} = 75 \times \frac{3070}{1020} = 225.7 \text{ cm.}$$

Pressure at a point. A point has no area, and as pressure is force *per unit area*, it may seem strange to speak of pressure at a point. We get over the difficulty by considering an extremely small area including the point in question, and we then define **pressure at a point** as the pressure on an extremely small area surrounding that point. (Cf. velocity at a given instant, p. 4.)

Specific gravity by U-tube. On p. 77 of *Junior Physics* we considered a 'U-tube' method of finding specific gravity (Fig. 10/4). BFD is mercury, AB is the

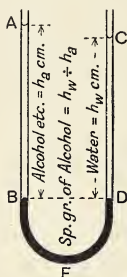


FIG. 10/4

liquid whose specific gravity is being found (say alcohol), and CD is water, the quantity being so adjusted that B and D are at the same level. If AB is h_a cm. and CD is h_w cm., it can be shown that the required specific gravity is $h_w \div h_a$.

For the reason already given in our earlier book, the U-tube method is not very accurate. Much better results are obtained by using apparatus similar to that designed early in the 19th century by Robert Hare, who was Professor of Chemistry in the University of Philadelphia.

Fig. 10/5 represents a determination of the specific gravity of alcohol, contained in the beaker A, while water is contained in W. The liquids are sucked up at C, after which the clip is closed. The specific gravity is equal to *height of water column (EM) \div height of alcohol column (DL)*.

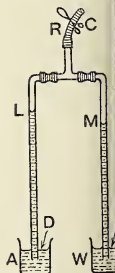


FIG. 10/5

Measuring gas pressure. For measuring small pressures such as that of the domestic gas supply, a U-tube containing water or oil may be used. The gas is connected as shown in Fig. 10/6, and the height h is measured. The required pressure is that experienced at A, and this is, of course, equal to the product hs , where s is the specific gravity of the liquid used.

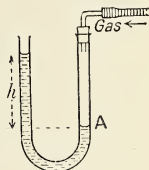


FIG. 10/6

The siphon. Suppose the tube ACDB in Fig. 10/7 is full of water. The water will not move if the pressure in one direction is greater than that in the other.

Now, if P = atmospheric pressure, the upward pressure at A is $P - x$ (where x is the pressure due to the column of liquid AC).

Similarly, upward pressure at B is $P - y$.

$$\therefore (\text{pressure at A}) - (\text{pressure at B}) = (P - x) - (P - y) = y - x.$$

Thus the pressure at A, tending to drive the liquid from left to right, is greater than that at B, tending to drive it from right to left; and so the liquid will flow from left to right.

When, owing to the transfer of liquid, both quantities are at the same level (L and M), $y - x = 0$. The pressure in the two directions is then equal, and liquid ceases to flow. Notice that if x exceeds the height of a barometer made of the liquid in question (34 ft. in the case of water), the connecting tube will not remain full, and there will be no siphon action.

To sum up, the conditions for siphon action are (see Fig. 10/7):

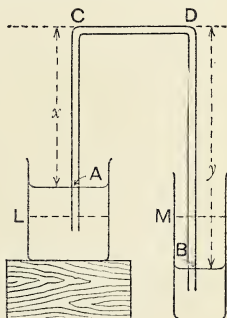


FIG. 10/7

- (i) Tube ACDB must be full of liquid.
- (ii) Height CA must not exceed the height of a barometer made of the liquid in question.
- (iii) B (i.e. surface of liquid in receiving vessel) must be at a lower level than A (surface in supplying vessel).

A siphon is often used for emptying a vessel which cannot conveniently be tipped up, or for removing an upper layer of liquid when it is desired to leave the lower layer undisturbed.

In a common type of lavatory flush the exit pipe is covered by a heavy iron bell. On raising this and allowing it to fall, water is pushed into the exit pipe, causing a siphon action to set in which continues until the tank is empty.

The tank is filled from the inflow pipe T, which is open and closed by the movement of a lever, operated by the floating ball B.

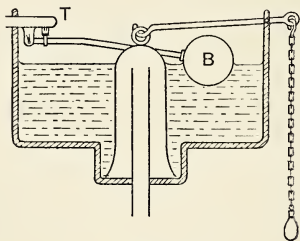


FIG. 10/8

Transmission of pressure. In Fig. 10/9 A is a cylinder of cross-section, say, 100 sq. in. connected by a pipe C with a much narrower cylinder B, of cross-section say 1 sq. in. Cylinders and pipe are completely full of water, A is closed at the top by the disc E, and B by the plunger D. Suppose we apply a downward force of 5 lb.wt. to the plunger, what will be the effect on E, the upper surface of A?

Pressure is transmitted equally in all directions through a fluid, and therefore E will experience the same pressure that was applied to B—viz. 5 lb. per sq. in. But E has an area of 100 sq. in., and so it will experience a total thrust of $100 \times 5 = 500$ lb.wt.

Let us suppose the cylindrical sides of A are continued upwards as shown by the dotted lines until they are terminated by the fixed cover FG. Let E have an up-and-down sliding movement. By pressing D with a given force we can move

ing 100 times that force to bear on E, and so we can compress any object placed between E and FG.

Now, suppose that D is pushed down 1 ft. E rises only $\frac{1}{100}$ th ft., because, though the volume of water entering A is equal to that leaving B, it has to spread itself over 100 times the area. We therefore have (if the force applied to A is 5 lb.wt.)—

$$\text{work put into machine} = 5 \times 1 = 5 \text{ ft.lb.}$$

$$\text{work got out of machine} = 500 \times \frac{1}{100} = 5 \text{ ft.lb.}$$

an example of the Principle of Work.

The mechanical advantage of the ideal machine is, of course, $500 \div 5 = 100$.

Our machine as indicated so far is a somewhat theoretical arrangement, but a few improvements soon suggest themselves. We need a valve in the pipe C, so that, though water can be pushed from B to A, none can flow

back. Again, when we draw D back in preparation for another push, it will be necessary that B should fill up with water from a valve-controlled inlet. These and other modifications were made by the Yorkshireman Joseph Bramah in 1795, the machine in its final form being known as the *Bramah Press* or

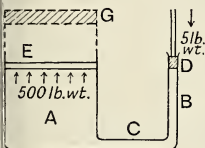


FIG. 10/9

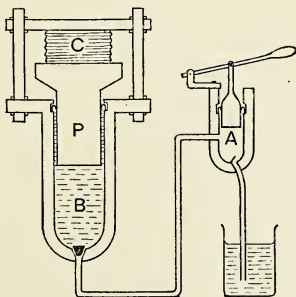
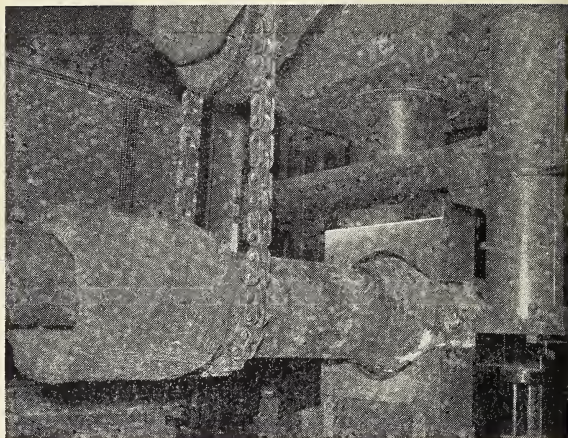


FIG. 10/10



By courtesy of English Steel Corporation, Ltd.

ELECTRO-HYDRAULIC PRESS AT WORK

hydraulic press. Fig. 10/10 will give a fair idea of its general construction, though many details are omitted.

One of the inventor's difficulties was that when pressure was being applied, water tended to escape past the pistons. If he tried to avoid this by making the pistons fit very tightly, it became nearly impossible to move them.

He got over this finally by making a groove in the cylinder, into which fitted a circular leather band, folded over so that its cross-section was the shape of an inverted U. We can see that as the cylinder advances, water will be pushed into this 'U' shape, making the movable flap press hard against the cylinder and so completely preventing escape of water.

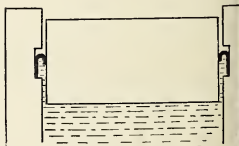


FIG. 10/11

The Bramah Press is used for compressing bales of cotton.

squeezing oil out of seeds, shaping steel ingots and many other purposes. The press in the illustration is being used to forge the end of a boiler drum, which it is squeezing with a force of 7000 tons weight.

The *hydraulic jack* used for raising motor-cars is very similar in principle to the press just described, but it is operated with oil instead of water. In Fig. 10/10 we must imagine the car to be in the position C, the frame round which has been removed.

The cylinders A and B are connected by an additional pipe not present in the ordinary hydraulic press, and so not shown in the figure. This pipe includes a tap, and when we wish to lower the car we open the tap, whereupon oil flows back into the small cylinder, and the piston P falls.

B. Gases

Here we shall take up and extend, quite briefly, the chief points with which we dealt in our earlier book.

We saw there that the pressure of the atmosphere is measured by the barometer, and we referred to a 'rather curious defect' commonly found in this instrument. Let us see what this is. Suppose our instrument is as shown in Fig. 10/12, the surface of the mercury in the tube being at C and that of the mercury in the dish at D. CD is the 'height of the barometer,' and the scale by which we measure height should have its '0' mark at D.

Now, suppose the mercury rises owing to an increase in pressure. More mercury in the tube means less in the dish, the level of D falls. Hence to get a true reading we should be obliged to lower our scale slightly. This is much too inconvenient for practice, and we have to content ourselves with an inaccurate reading.

To get over the difficulty, the mercury reservoir is made to consist of a wash-leather bag, the bottom of which can be moved up or down by turning a screw (Fig. 10/13). This is done until the mercury surface just makes contact with the tip of an ivory pin, and it is then in the correct zero position. For its inventor, this form of the instrument is known as the

Fortin barometer. It is nearly always fitted with a vernier scale, so that a more exact reading can be taken. This scale (PQ, Fig. 10/13) can be moved up or down by turning screw S. This is done until its base is level with the surface of the mercury, and the reading is then taken. In Fig. 10/13 it would be 29.37 in.¹

It is worth while noting that in meteorological measurements



FIG. 10/12

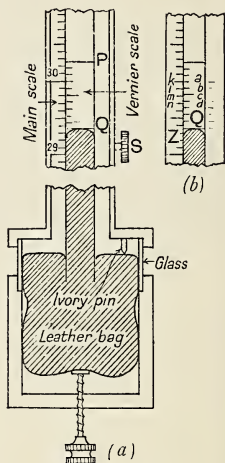


FIG. 10/13

pressure is expressed not in inches or centimetres of mercury but in *millibars*. (A millibar = 1/1000th of a bar, which is a pressure of 10^6 dynes per sq. cm.)

¹ The vernier scale PQ is 9/10ths of an inch long, and is divided into 10 equal parts, so that the length of each part is $9/10 \div 10$, or 9/100 of an inch.

To read the vernier, we find that division of it which coincides most nearly (or exactly) with a 'main scale' division. In Fig. 10/13 this is *a*, which coincides with *k*. Now, since *ab* = 9 hundredths of an inch, while *kl* = 10 hundredths (one-tenth), it is clear that *bl* = 1 hundredth, or 0.01. Similarly, *cm* = 0.02, *dn* = 0.03, and so on. Working down in this way, we find that *QZ* = 0.07. Thus the required reading is $29.3 + 0.07 = 29.37$ in.

Example 3. If the relative density of mercury = 13.6, and 1 gm. = 981 dynes, express 'normal pressure' (76 cm.) in millibars.

the height of the barometer is the same whatever the cross-section, suppose the latter to be 1 sq. cm. You should now be able to go on for yourself. The answer is **1014**.

Let us briefly recall **Boyle's Law**, which states that *If a given mass of gas is kept at constant temperature, its volume varies inversely as the pressure.*

Fig. 10/14 shows a convenient form of apparatus by which it may be demonstrated. The given mass of gas is enclosed in the part CD, the pressure is the atmospheric pressure (as shown by the barometer) *plus* or *minus* the difference between the levels at D and E.

Q. Which is it—plus or minus—in the figure ?]

Example 4. A strong glass tube 30 cm. long closed at one end is lowered to the sea-bottom end downwards, and then drawn up. By a similar method it is found that the sea-water must fill the tube to within 1 cm. of the closed end. At the depth of the sea at this point, assuming the temperature is the same as at the surface. Atmospheric pressure = 76 cm. of mercury. Density of mercury = 13.6 gm. per c.c., and of sea-water = 1.02 gm. per c.c.)

Density of mercury : density of sea-water = 13.6 : 1.02.

A barometer made of sea-water would have

height of $76 \times \frac{13.6}{1.02}$ cm. = 1013 cm.

Now, the air in the tube has been compressed to 1/30th of volume.

It has been under a pressure of 30 atmospheres.

The air pressure on sea surface accounts for 1 atmosphere.

The sea-water accounts for 29 atmospheres.

To produce 1 atmosphere pressure, we need 1013 cm. of sea-

water. For 29 atmospheres we need 1013×29 cm. = **294 metres**.

Use of pressure in gases. Suppose a man A is holding a heavy iron plate in front of himself, while another man B

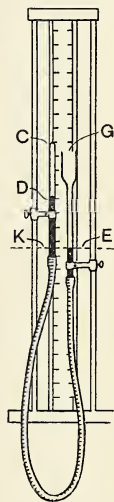


FIG. 10/14

hurls golf balls at him. He would feel the arrival of each as a single sharp blow. If B obtained a number of assistants A would, of course, feel a quick succession of sharp blows.

Suppose B were able to get so many assistants that his shield was receiving 40 or 50 blows per second. This frequency is so great that A would no longer be able to distinguish separate blows, but he would feel a fairly steady pressure against his shield.

Now, a great deal of evidence points to the fact that a gas such as air is not 'all in one piece,' but consists of an enormous number of rapidly moving particles called molecules, each being separated from one another quite as definitely as a golf ball (in the volley of golf balls) is separated from another. Small as these molecules are, a great deal is known about their size, speed, etc. Thus the number present in 1 cu. cm. of air at N.T.P. (i.e. at 0° C. and a pressure of 76 cm. Hg) is 2.7×10^{19} . That does not convey very much, so let us get a rough idea in another way. The population of the world is somewhere about 2000 millions. If the 2.7×10^{19} molecules were replaced by pound notes, there would be more than enough to give 2000 million pound notes to every one of the 2000 million people on the surface of the earth.

The molecules are continually flying about, continually hitting one another and rebounding, each molecule colliding some millions of times per second with its companions. The speed of these flying particles is very variable, but in the case of air at ordinary temperature it is not far short of the speed of a rifle bullet. (Air is a mixture of nitrogen molecules and oxygen molecules, with smaller numbers of some others.)

Naturally these molecules are continually hitting the sides of the containing vessel and then rebounding. It is this very rapid succession of impacts that produces pressure, just as much as it was produced by the rain of golf balls in the preceding illustration.

Now let us crowd these millions of molecules into a small space—we did this in our 'Boyle's Law' experiment, by making the mercury rise in the closed limb of the tube (*Junior Physics*, p. 100). They will now hit the sides of the vessel much more frequently, and therefore the pressure will be increased. Mathematicians can prove that if a gas is made up of molecules

es, as just described, the pressure ought to vary inversely the volume. This is, of course, exactly what we found in experiment.

If you have a dented ping-pong ball and warm it in front the fire, it straightens out, owing to an increase in the internal pressure. What has happened from the molecular point of view?

The application of heat has increased the speed of the molecules, and they now pound the inner surface of the ping-pong ball with greater force—i.e. greater pressure is exerted. As in the statement of Boyle's Law it is always specified that the temperature must be constant; otherwise we should have a change of pressure, even if the volume remained constant.

Questions

A

1. By how much would the pressure at a point 20 cm. below surface of mercury exceed that at the surface? (Sp. gr. of mercury = 13.6).

2. Using Hare's apparatus with water and alcohol, the water column is found to be 16.6 in. high and the alcohol column 20 in. Find the sp. gr. of alcohol.

3. In some experiments to find the specific gravity of paraffin by means of Hare's apparatus, the following results were obtained :—

Height of water column (cm.)	25.0	30.0	35.0	40.0	45.0	50.0
Height of oil column (cm.)	31.0	37.0	44.0	49.8	55.9	62.4

Find the specific gravity in each case, and also obtain the average. Work to 3 significant figures throughout.)

Find the pressure in lb. per sq. in. at the bottom of a lake 10 ft. deep. (1 cu. ft. of water weighs 62.3 lb.; atmospheric pressure 14.7 lb. per sq. in.)

A cylindrical tank of 'static' water is 9 ft. deep, and the area of the base is 1120 sq. ft. Find (i) the pressure at a point at the bottom in lb. per sq. ft., (ii) the total thrust on the bottom of the tank. Assume that 1 cu. ft. of water weighs 62.4 lb., and neglect atmospheric pressure.

B

1. The Dead Sea has a maximum depth of 1284 ft. What would be the pressure in lb. per sq. in. at this depth, given that the mean specific gravity of the water is 1.17? What would the pressure at this depth in fresh water? (Neglect atmospheric pressure. 1 cu. ft. of fresh water weighs 62.4 lb.)

2. Calculate the total thrust on a square plate of side 12 cm. immersed horizontally 20 cm. below the surface of a liquid of density 0.8 gm. per c.c.

3. Find the thrust exerted on the bottom of a vessel of uniform cross-section 50 sq. cm. by water which fills it to a depth of 20 cm. Find the additional thrust on the bottom of the vessel when a cubical block of wood of volume 50 c.c. and sp. gr. = 0.5 floats freely on the surface. *O.C.**

4. A cubical tank of 1 metre edge is filled with water. What is the resultant thrust on the base and in what units is it measured? *O.C.**

5. A tank of 'static' water is 40 ft. long, 20 ft. broad and 4 ft. deep. Calculate the total thrust on (a) the bottom of the tank, (b) one of the shorter sides. (The mass of 1 cu. ft. of water is 62.5 lb.) *Lond.**

[Hint. Find the average pressure on the side, i.e. the pressure at a depth of 2 ft., and then multiply by the area to obtain the total thrust.]

6. Calculate the total thrust on a square plate of side 10 cm. immersed horizontally 15 cm. below the surface of a liquid of density 0.8 gm. per c.c. *O.C.**

7. Describe a method you would employ to measure the pressure of the gas supply in the laboratory.

If the gas pressure is read as 5 in. of water, while the barometer reads 30 in. of mercury, calculate the total pressure of the gas in lb.wt. per sq. in. (Specific gravity of mercury = 13.6, and 1 cu. ft. of water weighs 62.4 lb.) *C.W.B.*

8. Describe and explain fully the action of the siphon in transferring a liquid from one vessel to another.

Two cylindrical beakers of the same height, but of cross-sectional areas 30 and 100 sq. cm. respectively, stand side by side on a table. The small beaker is filled with water, and the larger one is empty. What fraction of the water will it be possible to transfer from the smaller to the larger beaker by means of a siphon? *Camb.*

9. The larger piston of a hydraulic press is 12 in. in diameter and the smaller piston $\frac{1}{2}$ in. in diameter. What force will the larger piston exert when the smaller has an effort of 30 lb. applied to it? *Lond.**

0. In a hydraulic press, the diameter of the larger plunger is 5 cm. and the diameter of the pump plunger is 5 cm. Calculate the force exerted by the ram when a force of 2 kgm.wt. is applied to the smaller plunger. *O.C.**
1. Describe and explain the action of a Fortin barometer. What errors will be introduced (a) if the barometer is not vertical, (b) if the mercury is damp? *O.C.**
2. Explain what is meant by atmospheric pressure, and describe an instrument by which it can be measured. Indicate why the reading of a mercury barometer is independent of the cross-sectional area of the column.
A vertical barometer tube 100 cm. in length, and dipping into mercury, contains a small quantity of air. When the open end is 5 cm. below the surface of the mercury, the meniscus is 30 cm. from the upper closed end. When the tube is pushed into the mercury so that the open end is 35 cm. below the surface, the meniscus is 20 cm. from the closed end. Calculate the atmospheric pressure. *O.C.*
3. Some dry air is enclosed by a pellet of mercury 10 cm. in length in a uniform capillary tube, which is sealed at one end and open to the atmosphere at the other end. When the tube is placed vertically with the open end uppermost, the length of the air column is 13.3 cm. When the tube is inverted so that the closed end is uppermost, the length of the air column is 17 cm. Calculate the pressure of the atmosphere. *Camb.**
4. A barometer contains some air above the mercury column in consequence reads 72 cm. when the true barometric pressure is 75 cm. of mercury. The top of the tube is 82 cm. above the mercury in the trough. What is the true pressure when the reading is 68 cm., the tube being adjusted so that the top is still 82 cm. above the mercury in the trough? *Dur.**
5. Calculate the atmospheric pressure in dynes per sq. cm. when the mercury barometer stands at 77 cm. (Sp. gr. of mercury = 13.6, and $g = 981 \text{ cm./sec.}^2$)
6. Calculate the pressure in *bars* if the barometric height is 760 mm. of mercury. (Density of mercury = 13.6 gm. per c.c.; $g = 981 \text{ cm./sec.}^2$)

CHAPTER 11

THE PRINCIPLE OF ARCHIMEDES (RESUMED)

THE ancient Greeks pursued scientific studies to a considerable extent, but were not much given to exact measurement. As a result, they often went wrong when dealing with a question in which *quantity* was involved. We have seen, for instance, that they believed (quite wrongly) that the rate at which a body falls to the ground is proportional to its weight.

The 'Principle of Archimedes' is an exception, probably because it can be arrived at simply by *thinking*—and the Greeks were great thinkers. In Chapters 10 and 11 of our earlier book we have already dealt with it at some length. In the present chapter we shall concern ourselves for the most part with some of its more difficult applications.

The Principle may be stated thus: *If a body is partly or completely surrounded by a fluid, it will experience an upward thrust equal to the weight of fluid it is displacing.*

Comparing this statement with that given on p. 108 of our earlier book, two differences will be noticed. (i) The insertion of the words 'partly or completely.' This is done to include the case of floating bodies. (ii) The word 'fluid' has been used instead of 'liquid.' This term (Lat. *fluere* = to flow) includes gases as well as liquids, for we shall see presently that the Principle holds for gases. Thus the statement given above may be regarded as the Principle of Archimedes in its most general form.

In the earlier book we saw that by weighing a body in air and then in water we could find its specific gravity. Let us now consider the more awkward case that arises when a substance is soluble in water—a crystal of rock-salt, for instance.

Here we must choose some liquid in which the substance is insoluble. In the case mentioned, turpentine may be used. The method will be clear from an example.

Example 1. A piece of rock salt weighs 13.24 gm. in air, and 7.75 gm. in turpentine of specific gravity 0.90. Find the specific gravity of the rock-salt.

Wt. in air	13.24 gm.
„ „ turpentine	7.75 „
∴ Loss of wt.	5.49 „
∴ wt. of turpentine displaced	=	5.49 gm.			
But sp. gr. of turpentine	=	0.90			
∴ vol. of turpentine displaced	=	$\frac{5.49}{0.9}$	=	6.10 c.c.	
∴ vol. of rock salt	=	6.10 c.c.			
But wt. of rock salt	=	13.24 gm.			
∴ sp. gr. „ „	=	$\frac{13.24}{6.10}$	=	2.17.	

N.B. (i) If the specific gravity of the liquid is not given must be found, e.g. by use of the density bottle.

(ii) In some cases the most suitable liquid is a saturated solution of the given solid. In this, of course, the solid is soluble.

Specific gravity of floating body. We can also find the specific gravity of a small floating body, e.g. a cork, by means of the Principle of Archimedes. A suitable 'sinker' is obtained (e.g. a glass stopper, piece of lead, etc.) and we then find (i) weight of sinker in air, (ii) weight of sinker in water, (iii) weight of cork + sinker in air, (iv) weight of cork + sinker in water. The result worked out as shown below.

Example 2. A piece of cork weighs 1.48 gm. It is attached to a 20 gm. brass weight and the combination weighs 13.18 gm. in water. The brass weight alone weighs 17.62 gm. in water. Find the specific gravity of the cork.

The volume of a body in c.c. is numerically equal to the 'loss of wt.' (in water) in grams (*Junior Physics*, p. 111).

∴ Vol. of brass weight = $20 - 17.62 = 2.38$ c.c.

Again,

wt. of brass + cork, in air = $20 + 1.48 = 21.48$ gm.

wt. of brass + cork, in water = 13.18 gm.

∴ vol. of brass + cork = 8.30 c.c.

But vol. of brass = 2.38 c.c.

∴ vol. of cork = 5.92

∴ sp. gr. of cork = $\frac{\text{wt.}}{\text{vol.}} = \frac{1.48}{5.92} = 0.25.$

Principle applied to floating body. You will remember the apparatus by which we weighed a body in water—balance small wooden stool, etc. (Fig. 11/1).

Now if we were to weigh a cork, piece of wood, etc., in this way the cotton by which we were suspending it would be slack, and the 'weight in water' would be zero. Therefore in the case of a floating body, the *loss of weight*, or upthrust would be $(\text{wt. in air}) - (\text{wt. in water}) = (\text{wt. in air}) - \text{zero}$
 $= \text{wt. in air}.$

In short, for a floating body, $\text{loss of wt.} = \text{wt. in air}.$

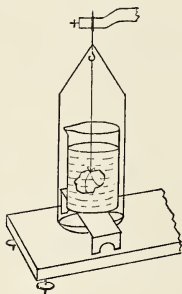


FIG. 11/1

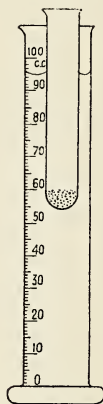


FIG. 11/2

Thus the statement 'loss of weight equals weight of liquid displaced' becomes 'the weight of a floating body is equal to the weight of liquid displaced.' This statement is often known as the Principle of Flotation. It is really a special case of the Principle of Archimedes—the Principle applied to floating bodies.

It is worth while recalling how we tested the truth of the Principle of Flotation by experiment. As our floating body we took a test-tube containing some sand or lead shot, and weighed it. We then floated it in water contained in a measuring cylinder (Fig. 11/2), and by taking the reading before

was put in, from the reading afterwards, we found the *volume* of water displaced, in cubic centimetres, which, of course, gave the *weight* of water displaced, in grams. This found to be equal to the weight of the floating body. Here a fairly difficult example which involves this principle.

Example 3. *A rectangular wooden raft measuring 2 m. \times 1.5 m. \times 3 m. is being used for salvaging lead from a damaged vessel. It is decided to take as the 'maximum safe load' half the weight of lead which would just produce submergence of both raft and lead. If the specific gravity of the wood is 0.7, of lead 11.4, and of sea-water 1.03, find the maximum safe load.*

In this problem it will be best to work in cubic decimetres. 1 cu. dm. of water (1000 c.c.) weighs 1000 gm. = 1 kgm.

Vol. of raft = $20 \times 15 \times 3$ cu. dm.

$$= 900 \text{ cu. dm.}$$

$$\text{wt.} = 900 \times 0.7 = 630 \text{ kg.}$$

Suppose wt. of lead required for complete submergence = x kgm.

$$\text{Vol. of lead} = \frac{x}{11.4} \text{ cu. dm.}$$

$$\text{Total vol. of raft and lead} = 900 + \frac{x}{11.4} \text{ cu. dm.}$$

When just on the point of submerging, vol. of sea-water displaced

$$= 900 + \frac{x}{11.4} \text{ cu. dm.}$$

$$\text{wt. of sea-water displaced} = \left(900 + \frac{x}{11.4}\right) \times 1.03 \text{ kgm.}$$

But this = wt. of floating body (raft + lead)

$$\left(900 + \frac{x}{11.4}\right) 1.03 = 630 + x.$$

Solving, we get $x = 361.3$ kgm.

$$\text{'maximum safe load'} = \frac{1}{2} \text{ of } 361.3 = \mathbf{180.6 \text{ kgm.}}$$

Balloons and airships. Applying the Principle of Archimedes in the case of a body such as a balloon surrounded by air, we see that the upthrust should be equal to the weight of air displaced. A proof of this is given on pp. 107–8 of our earlier book, except that the brass weight mentioned there should be replaced by a balloon, the liquid being replaced by air.

Here we will give a very brief résumé of the proof.

Imagine A to be held captive by a rope BC (Fig. 11/3). The tension in the rope indicates an upthrust from the

surrounding air (the upper air is pressing downwards D_1, D_2, D_3 , etc., and the lower air is pressing upwards E_1, E_2, E_3 , etc. The 'E's predominate because pressure greater as we come nearer to the earth; hence the resultant pressure is upwards.)

How great is this upthrust?

If the balloon disappeared and its place were taken by the air that was originally in that position, the latter would move neither upwards nor downwards. But this air obviously

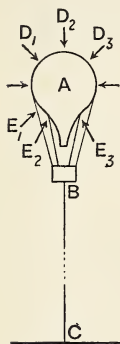


FIG. 11/3

has *weight*, which in the absence of other forces would cause it to move downwards. Therefore the outside forces acting on it are on balance,¹ equal to this weight, and are acting upwards.

These outside forces are just the same whether the space A is occupied by air or by the balloon. The balloon therefore is subject to an upward force equal to the weight of the air which it displaces.

It is now easy to see why a balloon rises if it is filled with some light gas such as hydrogen. Suppose the hydrogen weighs 1 cwt. and an equal volume of air weighs $14\frac{1}{2}$ cwt. Suppose also that the envelope of the balloon, ballast, basket, etc., together weigh 4 cwt. The total weight (allowing for the hydrogen) is 5 cwt. The upthrust is $14\frac{1}{2}$ cwt. Therefore the lifting power of the balloon would be $9\frac{1}{2}$ cwt.

It is worth while repeating this calculation for helium (which is non-inflammable). Helium is twice as dense as hydrogen and so would weigh 2 cwt. The lifting power comes out $8\frac{1}{2}$ cwt.—not so very much less than before. The chief objection to helium is its comparative rarity, for the only considerable supplies are found in certain oil-wells of the United States. America will not sell the gas to foreign countries, because she knows that airships may be used for other purposes than those of peaceful transport.

¹ I.e. allowing for the fact that some of the forces act from above and downwards, but are more than balanced by those acting from below and upwards.

The barrage balloon, all too familiar during the war, is made with a double skin. It is filled with hydrogen, but under the skin there is a scoop through which air can be admitted into the space between the two skins. When the balloon is freshly filled with hydrogen, these are pressed together and all the air is squeezed out. As it rises into regions of lower pressure the hydrogen expands, the surplus escaping through a valve into the atmosphere. When brought to lower altitudes the increased pressure would cause the balloon to be rather floppy and shapeless, but this is prevented by the entry of air through the scoop.

The fins at the rear serve to keep the balloon pointing into the wind. They are filled with air, which enters through scoops.

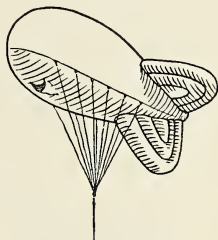
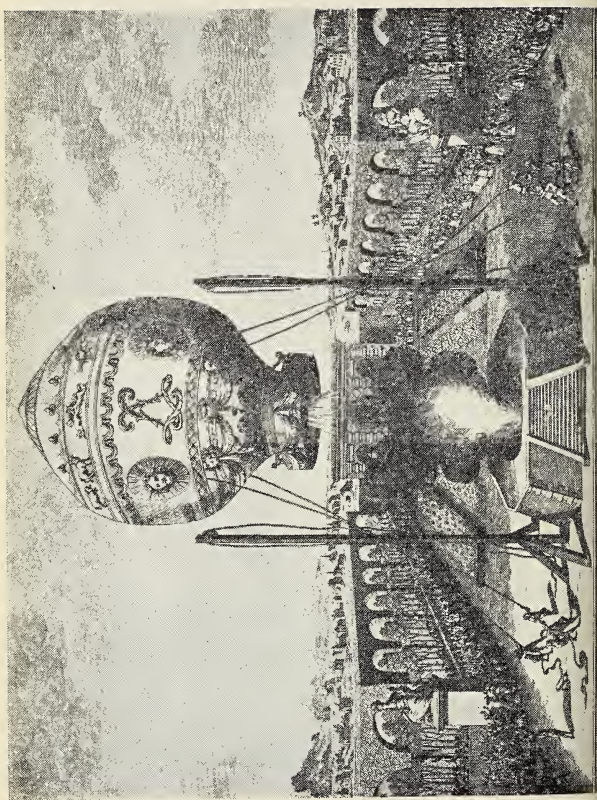


FIG. 11/4

Historical Note. Even in the most remote ages, man watched birds and dreamed of a day when he, too, would be able to fly through the air. Quite naturally he always thought of wings attached to his own body, and records of some very curious attempts at flight have been preserved. Thus early in the 16th century an Italian alchemist undertook to fly from the walls of Stirling Castle to France, and actually made the attempt, but was immediately in need of treatment for a broken leg. It appears that his wings, though made chiefly of eagles' feathers, which would naturally soar into the heavens, contained some derived from the ordinary hen, which tended to move downwards towards the dunghill. At that time, that was the explanation put forward by the unfortunate alchemist.

For real advance we must wait for the year 1783. The Montgolfiers had often noticed how clouds floated in the air, and they thought that if only they could enclose the cloud-like substance in a bag, the bag ought to float.

The most promising substance seemed to be smoke, so they experimented with a paper balloon with an open neck at the bottom, holding it over a smoky fire for some time, and then releasing it. We are not surprised that the experiment was



VUE ET PERSPECTIVE DU JARDIN DE M^r REVELLEON FABRIQUANT DE PAPIER.

successful, for of course the balloon would be filled with hot air which is less dense than cold air. In June 1783 they gave a great public demonstration, filling a linen globe 105 ft. in diameter with smoke from a fire fed with bundles of dried straw. When the balloon was released it soon rose to a great height, and after a 10 minutes' flight descended at a distance of $1\frac{1}{2}$ miles away. It was a great triumph!

Before the year was out men were actually making ascents in 'aerostatic balloons,' carrying a brazier of hot coals just under the neck of the balloon and adding fuel as required. At the same time, however, successful attempts were being made to ascend in balloons with hydrogen, and in December 1783 Dr. Charles¹ made a successful ascent in a balloon with a suspended car, substantially different from those with which we are familiar to-day. He ascended from the Tuileries, remained some time at an elevation of 2000 ft., and descended two days later at a little town 27 miles away. Three months later a balloonist succeeded in crossing the channel from England to France.

About 1850 men began trying to devise some means of constructing a 'dirigible' balloon (Fr. *diriger* = to direct), but it was not until the first year of the present century that any substantial success was achieved—actually by Count Zeppelin. Some progress has been made, but there have been some terrible disasters which serve as a reminder that many problems remain to be solved.

Questions

A

Fill in the blanks in the following table:—

Substance.	Wt. in air (gm.).	Wt. in water (gm.).	Vol. of body (c.c.).	Sp. gr.
Mercury . .	35.80	31.80
Lead . .	16.08	6.00
Aluminum . .	22.70	11.35
Iron	6.08	4.00
Platinum	5.21	2.70

¹ The discoverer of Charles' Law ('Heat,' p. 154).

2. A flat-bottomed test-tube, containing lead shot to make float vertically, sinks to a depth of 4.6 in. in water, 4 in. in brine and 5.5 in. in methylated spirit. Find the specific gravity of the brine and of methylated spirit.

3. To what depth would the test-tube of q. 2 sink in (i) paraffin oil, sp. gr. 0.80, (ii) sulphuric acid, sp. gr. 1.84, and (iii) glycerine, sp. gr. 1.26.

B

1. State *Archimedes' Principle* and explain how it is applied to find the specific gravity of a solid.

A piece of glass weighs 25 gm. in air, 15 gm. in water and 17 gm. in paraffin oil. A lump of copper sulphate weighs 16 gm. in water and 10 gm. in paraffin oil. Calculate the specific gravities of the paraffin oil and copper sulphate. *Lond.*

2. Define *specific gravity*. Describe a method of finding the specific gravity of a piece of cork.

A barge 120 ft. long and 20 ft. broad, whose sides are vertical, floats partially loaded in water. How many inches will it sink when 12.5 tons of cargo are added? (1 cu. ft. of water weighs 62.5 lb.) *Lond.*

3. A body (sp. gr. = 0.94) is fastened by a thread to the bottom of a vessel containing brine of sp. gr. = 1.04, so that the body is completely immersed in the liquid. Given that the tension in the thread is 2 gm.wt., find the volume of the body. *O.C.*

4. A cubical block of density 1.5 gm. per c.c. and weighing 96 gm. in air is suspended by a string so that it is completely submerged in a liquid of sp. gr. 0.75. What will be the tension in the string?

If the upper surface of the block is coated with cork of density 0.5 gm. per c.c. so that it floats with the upper surface of the cork in the surface of the liquid, what thickness of cork will be required? *O.C.**

5. A body which sinks in water is suspended from a spring balance and gradually lowered into water. State and explain the difference in the readings when the body is (a) just above the water surface, (b) partly submerged, (c) just wholly submerged, (d) a few inches below the surface. How could information obtained from these readings be used to find the specific gravity of the body?

What mass of copper wire must be wrapped round a piece of wax of mass 20 gm. and specific gravity 0.89, so that the wire will just sink in water? (Specific gravity of copper is 8.9) *Bris.*

6. Explain as fully as you can why an inflated rubber diaphragm is able to float with several men on board. State in general terms the physical principle involved.

A flying-boat rests on the surface of the sea and takes on a load of petrol. Find the additional volume of the hull submerged when it has received 150 cu. ft. of petrol of specific gravity 0.8. The specific gravity of sea-water is 1.03. *J.M.B.*

7. State the principle of Archimedes, and describe how you could verify it by experiment.

An airship has a gas capacity of 1,000,000 cu. ft. and weighs 1000 lb. when deflated. Calculate the force which is required to prevent it from rising when it is inflated with hydrogen.

(1 cu. ft. of air weighs 0.076 lb.; 1 cu. ft. of hydrogen weighs 0.053 lb.) *Camb.*

8. A cork ball on one pan of a beam balance is balanced by brass weights on the other pan. A bell-jar which can be evacuated by an air-pump is placed over the balance and sealed. Describe and explain what you would expect to observe as the air is gradually removed from the jar.

You are given a tangle of bare copper wire. Describe in detail how you would determine the specific gravity of the copper and the length of the wire without unravelling it. *Dur.*

9. Describe a hydrometer for measuring the specific gravity of a liquid and illustrate by a diagram showing the scale graduations. State the principle involved in its action.

A uniform glass tube provided with a loaded bulb at its lower end has a mass of 50 gm. It floats vertically in water with 8 cm. of stem exposed, and in a liquid of specific gravity 1.25 with 12 cm. exposed. Calculate the area of cross-section of the tube. *Lond.*

10. A solid wooden sphere of radius 6 cm. and specific gravity 0.8 is tied to the bottom of a vessel containing water by a string, so that the sphere is completely immersed. Calculate in dynes the tension in the string. What fraction of the volume of the sphere remains immersed if the string is cut and the sphere floats freely? *O.C.*

CHAPTER 12

SURFACE TENSION. VISCOSITY. OSMOTIC PRESSURE

A LARGE part of the present chapter will deal with what seem like contradictions to conclusions previously reached. We shall find, for instance, that (provided we do not take too much of it) solid steel may float on water, and that in certain cases water may flow uphill! These points, and others, will arise in connection with Surface Tension. Later in this chapter we shall be concerned with two other subjects, Viscosity and Osmosis. All three, strictly speaking, belong to a separate branch of Physics known as *Properties of Matter*.

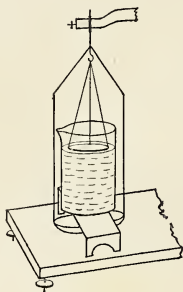


FIG. 12/1 (a)

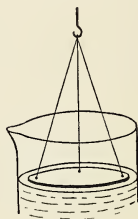


FIG. 12/1 (b)

and it is only as a matter of convenience that they have been put at the end of our 'Mechanics and Hydrostatics' section.

We will begin with a simple experiment in which a clean glass cover-plate is attached by means of wax to three pieces of thread. These are tied together at their upper ends and hung from the hook of a balance. Weights are placed in the other pan so as just to counterpoise the arrangement.

A beaker of water rests on a balance stool (as in specific gravity experiments), the amount of water being so adjusted that the plate just touches the surface. On raising the level

the balance the plate is found to 'stick,' and the force with which it sticks is measured by the extra weights that have to be placed in the other pan to cause the plate to break away. The under side of the plate is then found to be *wet*, showing that the extra weights were required not to separate glass from water, but to separate water from water. *The force with which parts of the same substance cling together is known as cohesion.*

If the experiment is repeated with mercury, the glass plate is found to come away *dry*, showing that mercury clings to mercury with more force than it clings to glass. *The force with which one substance clings to a different substance is known as adhesion.* Thus we speak of the adhesive force between glass and mercury. It has already been explained (p. 102) that substances are believed to be composed of extremely minute particles called molecules, and the experiments just mentioned prove that molecules attract one another, that the attraction of water molecules for glass molecules is greater than that of water molecules for one another; and that the attraction of mercury molecules for glass molecules is less than that of mercury molecules for one another.

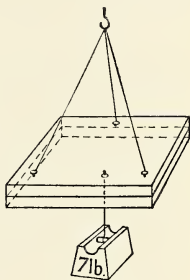


FIG. 12/2

It is evidently the force of cohesion which holds the different parts of a body together and causes it to resist 'breaking.' It is a force operating between a molecule and its neighbours, and acts only through extremely small distances. If it were otherwise, we could instantly repair broken crockery by pressing the separated parts together.

If care has been taken to make the surfaces of two slabs of glass very accurately plane, and these are pressed together, considerable force is required to separate them (Fig. 12/2)—the surfaces are so close that the force of cohesion is able to come into play. When we repair broken articles with glue, cement, etc., we are evidently making use of the force of adhesion. You can think out for yourself the reason why such substances have always to be applied when wet, or at least plastic.

Surface Tension. Think of a drop of water or other liquid. The molecules of which it is composed are all attracting on another, and so we should expect the drop to take the most compact form possible, which is a *sphere*.

Actually there are other forces—notably that of gravity—to consider in addition to the force of cohesion. The result is that the shape is not as a rule truly spherical. In the case of very small drops, however, the force of gravity is slight in comparison with that of cohesion, and the shape is very nearly spherical.

When drops of liquid (water, for instance) are lying on a surface such as a table-top, the force of *adhesion* is prominent and causes the drops to be non-spherical. It is true that very small drops of mercury lying on a table are spherical; but in this case adhesion is small in comparison with cohesion (as shown by the experiment of p. 117), and as the drops are very small, the force of gravity is slight.

Allowing for all disturbing factors, we may say that a drop of liquid tends to assume a spherical shape. It behaves as though it were covered with a thin elastic skin. Of course no such skin actually exists. What *does* exist is a force of attraction between one molecule and another. In what follows we shall usually employ the 'elastic skin' idea, because the effects that would be produced by such a skin are exactly the same as those produced by the attractive forces between molecules. The force produced by such an imaginary skin is known as **surface tension**, and may be defined as a force

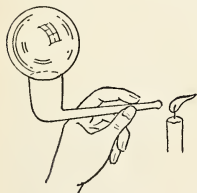


FIG. 12/3

at the surface of a liquid producing effects similar to those that would be caused if the surface consisted of a thin elastic membrane, but really due to the attractive forces which the molecules of the liquid exert upon each other.

Practical Effects. Surface tension is perhaps most obviously seen in a soap bubble. When one is blown on the end of a pipe, the surface tension causes contraction, and it is actually possible to blow out a candle flame by means of the air thus expelled (Fig. 12/3).

The steel of which a needle is made has a specific gravity

nearly 8. Yet a needle if laid gently down on water will float, supported by the surface tension.

Fig. 12/4 shows an end-on view of the floating needle.



FIG. 12/4

The material of which an umbrella is made is porous, and yet it does not let the water through. The threads are connected by a film of water, the tension of which is sufficient to resist penetration by the drops constantly falling on it. It is for the same reason that a tent gives protection against rain, and most campers are acquainted with the unfortunate results that follow if the film is broken by prodding.

Insects such as the water boatman are often seen 'skating' on the surface of water. They are, of course, supported by surface tension.

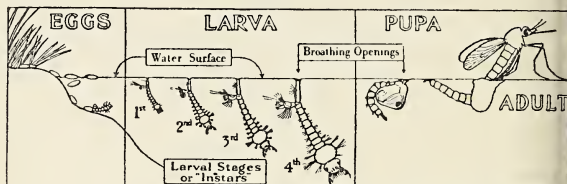
It is not surprising to find that surface tension varies greatly between one liquid and another. Thus the surface tension of water is greatly reduced by dissolving soap in it, and that is why a gardener so often adds soap to the liquid with which he is syringing rose-bushes, etc. High surface tension causes dew to collect in little balls on the surface of the leaves (this is easily observed after a slight shower). The soap solution, on the other hand, with its low surface tension spreads over the whole of the leaf, so that pests such as green-fly, etc., can be effectively dealt with.

Hot water has a lower surface tension than cold. Hence if two match-sticks are floated on water, parallel and about an inch apart, and if the liquid between them is touched with a wire, the match-sticks spring away from one another. The same thing happens if a drop of soap solution is placed between the water between the sticks. These results also you should be able to explain.

You have very likely read that malaria, which has been the cause of untold thousands of deaths in various parts of the world, is spread by certain mosquitoes, and so by killing the mosquito the disease is checked.

The mosquito lays its eggs on the surface of water, and they hatch out into *larvae*. The latter swim about under the surface, and by means of gills obtain some air from the water—not enough. They find it necessary to come to the surface

from time to time to rest and breathe, and they hang head downwards from the surface by means of a breathing tube which passes through the tail (Fig. 12/5), their weight being supported by surface tension.



By courtesy of the Director British Mosquito Control Institute

FIG. 12/5

By treating the water with paraffin a thin oil film is formed with a much lower surface tension than that of water. Hence the larvae find it nearly impossible to support themselves, and many of them drown.

The survivors develop into pupae, and from these the fully developed mosquitoes emerge. In the process, however, the mosquito is once again dependent on surface tension, and the presence of paraffin prevents this last essential change.

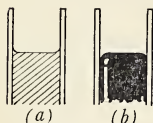


FIG. 12/6

When water touches glass the surface curves upwards (Fig. 12/6a). You must often have noticed this in a burette or measuring jar. In the case of mercury the curve is downwards (Fig. 12/6b). It can

be proved that these effects are due to surface tension, and we shall take them for granted in the course of the next paragraph.

Capillarity. Suppose we put a very narrow tube into water. We may suppose that at first the water 'finds its own level' (Fig. 12/7a—for clearness the width of the tube is greatly exaggerated), but the surface takes the curved form indicated. Think of a stretched skin lying along this curve. The middle portion will evidently be pulled up to the level of the tube edges (Fig. 12/7b), and then the edges will once more curve upwards (Fig. 12/7c). Once again the middle portion will be pulled up (Fig. 12/7d), and so on. The process will stop when

upward force due to surface tension is balanced by the downward force due to the weight of the column of liquid. We should expect that the thinner we make the tube the higher would be the column of liquid supported; and this is found to be the case.

In the case of mercury, surface tension would pull the column of liquid *downwards* to a point at which the upward pressure due to the higher level of the liquid outside the tube (Fig. 12/8) is balanced by the downward force due to surface tension.

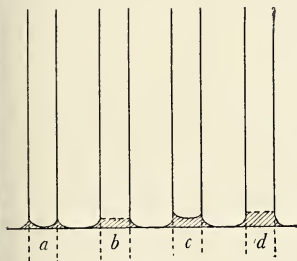


FIG. 12/7



FIG. 12/8

Narrow tubes such as are used for showing these effects of surface tension are called capillary tubes (Lat. *capillus* = a hair), and the phenomenon itself is known as capillarity. Instances of capillarity are extremely common in daily life.

We have an example in the soaking up of ink by blotting-paper, the pores in the paper serving as capillary tubes, and a very similar instance is seen when the corner of a lump of sugar is dipped into tea. An example with a very practical bearing is the continuous rising of oil in the wick of a lamp.

Gardeners sometimes quote a sort of proverb, 'the hoe is the best watering can.' The air spaces between grains of soil form irregular capillary tubes, up which water rises from below, and when it reaches the surface, evaporation takes place. In dry weather it is desirable to check this evaporation, and the gardener gets to work with his hoe. As a result, the water can now rise only to the level at which the broken capillary tubes are, some inches short of the surface.

Viscosity. Many liquids such as water, petrol and alcohol are 'thin' or limpid, i.e. they pour easily. Others such as treacle and lubricating oil are sticky or viscous, and can only be poured slowly. Viscosity is due to internal friction. Instead of one portion of a liquid slipping easily over another it passes over it with difficulty.

It is wrong to suppose that some liquids are viscous and others not. All liquids possess some degree of viscosity—is only a question of more or less. The following experiment will enable us to make a rough comparison of the viscosity of various liquids—say water, olive oil, methylated spirit and syrup (made by adding two or three parts of water to one of golden syrup).

A tube about 12 in. \times $\frac{1}{2}$ in. internal diameter may be used, closed at one end.¹ It is *nearly* filled with the liquid to be

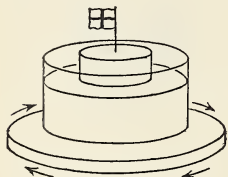


FIG. 12/9

examined, and corked—a fair-sized bubble of air being left between the liquid and cork. It is then inverted and the time required for the bubble of air to rise from the bottom of the tube to the top is noted. In a very viscous liquid such as olive oil the movement is extremely slow.

Another experiment to show that even a 'thin' liquid such as water possesses viscosity is described by Prof. Fleming in his 'Waves and Ripples.' A glass vessel half full of water is placed on a revolving table. Floating in the water is a round wooden disc carrying a paper flag mounted on a wire (Fig. 12/9).

When the table is turned slowly the disc is at first not affected. The glass vessel is, so to speak, slipping past the water contained in it. By and by, however, the flag begins to turn in the same direction as the table, and before long the glass vessel, water and disc are all moving round like one solid mass.

The experiment just described shows that there is a definite amount of friction between the outer layer of water and the containing vessel, and also between successive layers of water.

¹ Open tubes of these dimensions are often used for combustion experiments in chemistry and if corked up will serve very well.

man rowing on a river often makes use of this fact. Near the banks the current is weaker because of the friction offered by the banks and the bottom (the stream being shallower near the edge). Hence if he is rowing upstream he keeps near the edge, while if going downstream he keeps near the centre, so as to have the full advantage of the current.

Osmosis. In Fig. 12/10 a thistle funnel is shown with a piece of pig's bladder tied round it at A. It is surrounded by water in a beaker. Enough copper sulphate solution is poured into the funnel to bring the level up to C, that of the water outside.

On leaving the arrangement for some hours, it is found that the level of the copper sulphate solution steadily rises

(Fig. 12/10 it has risen to the point B). How do we explain this?

First of all, the membrane at A is what is known as *semi-permeable*, i.e. it will allow water to pass through it, but will not allow the passage of a solid dissolved substance, in this case copper sulphate. It acts as a sort of sieve.

This being the case, it looks as though water could pass quite freely from the beaker into the thistle funnel, and from the thistle funnel into the beaker. We imagine the dissolved substance as a sort of 'extra,' unable itself to pass through, but offering no opposition to the passage of the water in which it is dissolved.

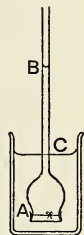


FIG. 12/10

But it is clear from the experimental result that water is passing *into* the funnel more freely than it is passing *out*. Why is it, when the liquid has reached a position such as B in the thistle funnel, higher than the level C in the beaker, that it does not drop till the level is the same on both sides?

Clearly it must be supported by pressure from below, and this pressure can only arise from the fact that copper sulphate is present in solution. We call it the *osmotic pressure* of the solution. Given a perfect semi-permeable membrane, liquid will continue to rise in the thistle funnel until the downward pressure due to the column of liquid is equal to the osmotic pressure of the solution.

In many ways osmotic pressure closely resembles gas

pressure. Indeed, it has been shown that, in general, x gram of a substance dissolved in v c.c. of water exert an osmotic pressure equal to the gaseous pressure that would be exerted if x grams of that substance were present as a gas with its volume adjusted to v c.c.; and there are other close analogies between osmotic pressure and gas pressure.

However, to go fully into these questions is rather beyond the scope of this book. Let us return to our copper sulphate experiment and summarise one or two practical conclusions.

First, the substance dissolved need not have been copper sulphate. We could have used sugar, common salt, or any other soluble substance. Similarly, there is a great choice of semi-permeable membranes.

Another point is that we need not have had pure water on one side of the membrane. We could have used copper sulphate solution on both sides, provided one solution was stronger than the other. In such a case the effect of osmotic pressure is to cause water to pass from the weaker solution to the stronger, thus tending to equalise their strengths.

Osmosis is very common in nature, as we shall see presently, and a semi-permeable membrane is often provided by the living cell walls of a plant. This may be shown by the following simple experiment.¹

Two large potatoes are peeled, and an end is cut off each so that a flat side is presented. Each potato is then hollowed out, leaving a wall about half an inch thick. One potato is boiled for three or four minutes, so as to kill the cells. Each

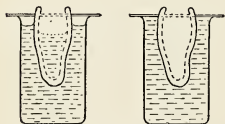


FIG. 12/11

potato is now supported in a beaker of water by means of a knitting needle passed through the upper end, and a good pinch of salt or cane sugar is placed in each. In a day or so a considerable quantity of water will have passed into the unboiled potato, but not into the boiled one, the reason being that the live cell walls have acted as a semi-permeable membrane.

The pressure due to osmosis is often very great. Thus the pressure set up by a 1% solution of saltpetre is enough

¹ From the *School Science Review* of Oct. 1935.

support a column of water over 100 ft. high. We begin to understand why osmotic pressure can account for the rising of the sap in trees.

Dried raisins swell when put into water. Inside the raisin is a solution of sugar, etc., and the skin forms a semi-permeable membrane, through which water from outside can pass. Raisins behave in the same way, and there are many other examples.

The action of various salts as a purgative—e.g. Epsom salts and Glauber salts—seems to depend largely on osmosis. The wall of the intestine acts as a semi-permeable membrane, and so water passes into the bowel from outside, thus serving to flush it.

If a doctor is bathing a wound with salt solution he likes to have the latter of 0.9% strength. The reason is that the cells in contact with the nerve endings are surrounded by a semi-permeable membrane, and contain a fluid having an osmotic pressure equal to that of a 0.9% salt solution. Hence when in contact with such a solution the cells neither gain nor lose liquid. Thus there is no change in the pressure exerted on the nerve endings, and consequently no sensation of smarting.

Questions

1. Distinguish clearly between *cohesion* and *adhesion*, and give examples of each.

2. Give reasons for the following:—

- (a) A needle may be floated on water.
- (b) Though an umbrella is made of porous material, it does not let the water through.
- (c) A gardener often adds soap solution to a liquid which he wishes to use for spraying.

3. If the end of a tube of very narrow bore is dipped into water, the water rises in the tube. How do you account for this? Mention two examples in daily life in which the same principle is at work.

4. It is sometimes found that 'tins' (made for containing goods such as cocoa, dried milk, etc.) hold water very well, but leak when paraffin oil is put in them. Why?



ANSWERS TO NUMERICAL EXERCISES

Answers to the odd-numbered questions only have been given.

Chapter 1, p. 9

A

(i) 60; (ii) 15; (iii) 90.

B

(i) 0; (ii) 48 ft./sec. 3. (a) $11\frac{1}{2}^\circ$ N. of E.; (b) 1 hr. 38 min.

Chapter 2, p. 24

A

(i) 5; (ii) 0.122. 3. (i) -32 ft./sec.²; (ii) 100 ft.

B

(a) 1392 ft.; (b) $38\frac{2}{3}$ ft./sec. 3. 100 ft.
(a) 17.7 sec.; (b) 6223 ft. 7. 144 ft.; 6 sec.

Chapter 3, p. 35

(a) 1.5 cm./sec.²; (b) 19.75 cm./sec. 3. 25.6 ft./sec.²

Chapter 4, p. 39

4 ft./sec.² 5. $\frac{Vt}{vT}$

Chapter 5, p. 49

A

(i) 21 ft.lb.; (ii) 1680 ft.lb.; (iii) 15 ft.lb. 3. $3\frac{1}{3}$.

B

(i) 144 f.p.s. units, 162 ft.lb.; (ii) 101.8 f.p.s. units, 81 ft.lb.
6720 ft.lb., 1.22.

39,270. 7. (a) 280 lb.wt.; (b) 513 lb.wt.

8. (a) 49,000 cm. gm.; (b) 98,000 c.g.s. units.

9. 1127 ft./sec. 11. (a) 240 ft./sec.; (b) 15 sec.

At ground level, P.E. = 0, K.E. = 900 ft.pdls.; at highest
t, P.E. = 900 ft.pdls., K.E. = 0; half-way up, P.E. = 450
ft.pdls., K.E. = 450 ft.pdls.

Chapter 6, p. 63

A

Lever, 50; 2; 1.98.

Block and tackle, 36; 4; 3. Inclined plane, 100; 0.8; 80.

3.6, 4, 90%.

B

28 lb.wt.; 4200 ft.lb. 3. 2 ft. 6 in.

1.40, 35%; 2.40, 60%; 2.71, 67.7%; 2.80, 69.9%; 2.80, 70.0%.

(a) 528; (b) 63.1%; (c) 396 ft.lb.

Chapter 7, p. 75

A

1. (i) 5 lb.wt. making \angle of 37° with greater force.
 (ii) 10 gm.wt. in direction of greater force.
 (iii) 50 gm.wt. in direction of original forces.
 (iv) 5.2 lb.wt. bisecting \angle between forces.
 (v) 18.5 lb.wt. bisecting \angle between forces.

B

1. 130.4 lb.wt. and 63.6 lb.wt. 3. $3\frac{1}{4}^\circ$ S. of E.; 8.5 lb.wt.
7. Tension in BC, 58.6 lb.wt.; in AC, 71.7 lb.wt.
9. 44.7 gm.wt., $26\frac{1}{2}^\circ$ N. of E. 11. 0.866 kgm., 0.500 kgm.

Chapter 8, p. 84

A

1. (i) $7\frac{1}{2}$ in.; (ii) 1.5 in. from middle point. 3. 30 gm.

B

1. 0.732. 3. (a) $2\frac{1}{2}$ lb.wt.; (b) $2\frac{1}{2}$ lb.wt.; (c) 10 ft.lb.
5. On a line joining the mid-points of the '1-2' and '3-4' sides and 3 in. from the latter.
7. 4.33 lb.wt. at 20° with the vertical; 12.76 lb.wt.
9. 36 cm. from rear of fuselage; 82 gm.

Chapter 9, p. 91

3. 0.33; (i) 130 gm; (ii) 150 gm.

Chapter 10, p. 103

A

1. 272 gm./sq.cm. 3. 0.806, 0.811, 0.795, 0.803, 0.805, 0.803.
5. 561.6; 280.8.

B

1. 651; 556. 3. 1000 gm.wt.; 30 gm.wt.
5. (a) 200,000 lb.wt. (= 89.3 tons wt.); (b) 10,000 lb. (= 89.3 tons wt.).
7. 14.9. 9. 17,280 lb.wt. (= 7.714 tons wt.). 13. 81.9 cm.
15. 1,027,300.

Chapter 11, p. 113

A

1. Top line, 4, 8.95; 2nd, 10.08, 2.68; 3rd, 20.70, 2.00; 10.08, 2.52; 5th, 14.07, 8.86.
3. (i) 5.75 in.; (ii) 2.50 in.; (iii) 3.65 in.

B

1. 0.80; 2.13. 3. 20 c.c. 5. 2.78 gm.
7. 40,700 lb.wt. (= 18.17 tons wt.). 9. 1.25 sq. cm.

PART II

HEAT, LIGHT AND SOUND

CHAPTER 13

HEAT AND TEMPERATURE. THERMOMETERS

Recapitulation. In Chapter 12 of *Junior Physics* we discussed the effects of heat on a body, and saw that these were mainly three.

- 1) There is a rise of temperature.
- 2) The body expands.
- 3) There may be a change of state, from solid to liquid or from liquid to vapour.

We know now that to these general rules there are exceptions. Thus when heat is applied to a beaker of melting ice there is no rise of temperature, and the same is true when heat boils water—its temperature remains at 100°C . As regards (2), alloys such as 'invar' have been prepared which undergo practically no change of volume when heated, and we have learnt that if water at a temperature between 0°C . and 4°C . is heated it actually contracts.

A change of state is no doubt always produced provided the necessary temperatures are available. In the sun, for instance, iron metals such as iron are present in the state of vapour.

When in the earlier book we learnt to distinguish between heat and temperature. We saw that the distinction is similar to that between *quantity* of water and *level* of water. If two barrels of water are made to communicate by means of a pipe, water flows from high level to low level quite irrespective of the *quantity* of water in each (Fig. 13/1). In the same way, if two bodies are placed in contact, heat will flow from the body at higher to that at lower temperature. When a blacksmith quenches a red-hot horse-shoe in a cistern of

water, heat flows from the horse-shoe to the water *not* because it contains more heat (as a matter of fact it contains much less), but because it is at a higher temperature.

Lastly, in our earlier work we saw how a thermometer is made.

- (1) A tube is prepared by blowing a bulb on one end. Near the other end the tube is drawn out to a narrow neck and a funnel containing mercury is attached.
- (2) By gently heating the bulb, some air is expelled, and on cooling, a little mercury enters. This mercury is then boiled to expel the remaining air, and on cooling, the bulb and tube are completely filled.

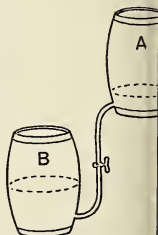


FIG. 13/1

- (3) The thermometer is now placed in a bath containing a suitable liquid, which is heated to a temperature slightly higher than the thermometer will ever be required to record. The tube is now sealed off at the neck.

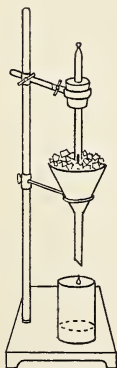


FIG. 13/2

- (4) The lower fixed point is marked (Fig. 13/2).

- (5) The higher fixed point is marked, special arrangements being made to secure that (a) the bulb and practically the whole stem shall be surrounded by steam and not simply by boiling water, (b) an outer jacket of steam shall serve to protect from draughts, etc. (Fig. 13/3).

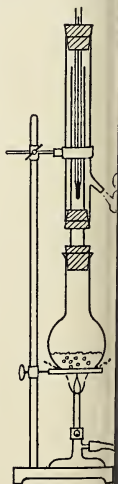


FIG. 13/3

We know now that the

temperature of boiling water depends on the pressure (*Junior Physics*, p. 169). In the neighbourhood of 100°C . an increase of 28 mm. in the pressure raises the boiling point 1°C . Thus in accurate work it would be necessary to read the barometer and make the proper allowance. For instance, if the barometer reading were 767 mm.—i.e. 7 mm. above normal—the boiling point would be $100\frac{1}{4}^{\circ}\text{C}$. (because 28 mm. raises the boiling point 1 degree, and 7 is a quarter of 28).

Self-Registering Thermometers. It is often necessary to know the lowest or highest temperature that has been reached during a certain interval. Thus a gardener may wish to know whether the temperature is falling so low in his greenhouse at night that he ought to start his heating apparatus. Hence 'maximum' and 'minimum' thermometers have come into use.

There are various forms of these, but we shall describe one which combines both maximum and minimum readings in a single instrument.

Max's Maximum and Minimum Thermometer.

This instrument is U-shaped, and the expanding liquid, consisting of alcohol, is contained in the upper part AB. The lower part of the U is occupied by mercury BCD, and covering D is a thin glass column of alcohol DE. The latter is only a minor part in the instrument—its purpose is to secure that the mercury surface at D shall be maintained under the same conditions as that at B.

Above the mercury in each limb there is a steel index (X) which just grips the glass by means of a weak spring. When the temperature is rising, the expanding alcohol pushes the mercury column BCD in front of it, and D in turn pushes the index X. When the temperature falls, the contraction of the alcohol at the other end of X will evidently record the highest temperature that has been reached. Similarly, the lower end of N will record the minimum temperature. When it is necessary to read the instrument, N and X are moved into contact with a mark K.

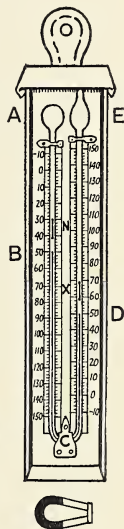


FIG. 13/4

their respective mercury surfaces by means of a magnet.

The instrument just described was invented by James S. about the year 1780.

The Clinical Thermometer. If a doctor were taking patient's temperature by means of an ordinary thermometer, he would have to read it while it was still *in position*, because the moment he removed it the reading would begin to fall. He therefore uses a special form of maximum thermometer known as a *clinical thermometer*. The instrument is shown to actual size in Fig. 13/5.

The liquid used is mercury. The special feature of the instrument is that at C, a little above the bulb, the tube is bent and constricted. Now, when the bulb is placed in the patient's mouth, the mercury has no difficulty in passing C, because it is driven on by the pressure from behind, and so the temperature is accurately recorded. When the thermometer is removed, the thread of mercury breaks off at C, the upper part remaining 'stranded' and continuing to record the highest temperature reached—i.e. the temperature of the patient. To re-set the instrument it must be shaken, when the upper portion of the mercury thread passes the constriction and joins the remainder.

It is worth while to consider the features of a good clinical thermometer, and how these are secured. The features in question are: (i) *accuracy*, (ii) *sensitiveness*, i.e. there should be a perceptible movement of the thread for a very small change of temperature such as $\frac{1}{10}^{\circ}$ F., (iii) *quickness of action*, (iv) *ease in reading*.

Accuracy is, of course, a matter of care in manufacture. The best clinical thermometers are not put on the market until they have been tested at the National Physical Laboratory.

Sensitiveness depends on the relation between the volume of the bulb and the bore of the tube. A large bulb and narrow bore both make for sensitiveness.



FIG.

an instrument which has to go into a patient's mouth is obviously not practicable to have a very *wide* bulb, but considerable volume is secured by making it *long*. The bore is very narrow, and the net result is that 1°F. is represented by about $\frac{1}{5}$ in. on the stem. It is thus quite easy to take readings to $\frac{1}{10}^{\circ}\text{F.}$ With such an open scale the *range* is of course quite small—usually from 95°F. to 110°F. —but this is all that the doctor requires. The temperature of the human body in health is about 98.4°F.

Swiftness of action. We want the whole of the mercury in the bulb to reach the patient's temperature as soon as possible, until this happens the thread will continue to rise. Obviously it is necessary that all the mercury should be as close as possible to the source of heat, this is favoured by having the bulb small and narrow, as already mentioned. Further, the glass of the bulb should be thin as is consistent with reasonable strength.

The quickest clinical thermometers are marked 'half-minute.' Actually, a somewhat longer time than that is needed for an accurate reading. The best ones are marked '3 minutes.'

Ease in reading. The thread of mercury is so thin that it would not be very

easy to take a reading, but the cylinder-shaped tube acts like a convex lens, greatly magnifying the width.

Transformation of scales. It is often necessary to transform centigrade readings to Fahrenheit, and vice versa. Consider the diagram (Fig. 13/6).

This reminds us that the interval between the freezing point of water and its boiling point is 100 degrees on the centigrade scale, but 180 degrees on the Fahrenheit scale.

Thus $100\text{ C. divisions} = 180\text{ F. divisions.}$

$$1\text{ C. division} = \frac{180}{100} = \frac{9}{5}\text{ F. divisions.}$$

$$1\text{ F. division} = \frac{5}{9}\text{ C. division}$$

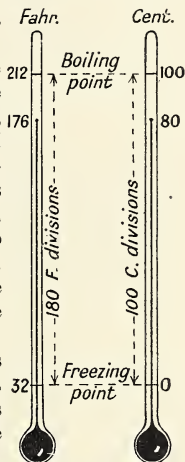


FIG. 13/6

Suppose we wish to reduce a temperature of 80° C. Fahrenheit.

$$1 \text{ C. division} = \frac{9}{5} \text{ F. divisions.}$$

$$\therefore 80 \text{ C. divisions} = \frac{9}{5} \times 80 = 144 \text{ F. divisions.}$$

Now 80° C. means 80 Centigrade divisions above freezing point. Hence the corresponding Fahrenheit reading is 144 Fahrenheit divisions *above freezing point*. But as freezing point is itself 32° F., the required temperature must $144 + 32 = 176^{\circ}$ F.

Next suppose we have to reduce a temperature of 68° to Centigrade.

Freezing point (Fahrenheit scale) is itself 32° , so 68° F. only $68 - 32$ or 36 degrees Fahrenheit *above freezing point*.

The corresponding number of Centigrade degrees *above freezing point* is $\frac{5}{9} \times 36 = 20$, and as Centigrade readings are reckoned from freezing point, the answer is 20° C.

By studying examples such as those just given, we see that the following general rules apply.

To bring Centigrade readings to Fahrenheit, multiply by $\frac{9}{5}$ and add 32.

To bring Fahrenheit readings to Centigrade, first subtract 32 and then multiply by $\frac{5}{9}$.

It is a good plan to remember that 212° F. = 100° C., and notice that to get from 212 to 100, we subtract 32 and multiply by $\frac{5}{9}$, while to get from 100 to 212 we multiply by $\frac{9}{5}$ (get 180), and then add 32. If you have to bring a Fahrenheit temperature to Centigrade and are uncertain as to whether you can remember the rule correctly, ask yourself, 'How do I get from 212° F. to 100° C.?' Similarly from Centigrade to Fahrenheit.

Example. *At what temperature would a Fahrenheit thermometer and a Centigrade thermometer give the same reading?*

Suppose the required reading is x° C. ($= x^{\circ}$ F.).

Now x° C. = $(\frac{9}{5}x + 32)^{\circ}$ F. by the rule given above.

But in this case the Centigrade and Fahrenheit temperatures are the same.

$$\therefore x = (\frac{9}{5}x + 32)$$

$$\therefore 5x = 9x + 160$$

$$\therefore x = -40.$$

$$\text{Ans. } -40^{\circ} \text{ C.} = -40^{\circ} \text{ F.}$$

Other Scales. Besides the two scales discussed above, there is a third scale much used in Germany, on which the freezing point of water is marked 0° and its boiling point 80° . It was invented by a French nobleman, the Seigneur Réaumur, about 1720, and is always known as the Réaumur scale.

80 seems a curious number to have chosen, but it arose in this way. Réaumur found that 1000 volumes of diluted spirits of wine (spirits diluted with one-fifth its volume of water) expanded 80 volumes between freezing point and boiling point, and it occurred to him that a very good 'degree' could be that amount by which the temperature of 1000 volumes of this diluted spirit must be raised in order to make it expand by 1 volume.

Of the three scales mentioned, the Centigrade is the simplest, and it is rather a pity that there should be any others in ordinary use. The multiplicity is confusing and often makes calculations necessary. However, the situation might have been much worse, for J. H. Pambert writing in 1779 gives a list of no less than nineteen! Sixteen of them have long since been dead and buried, and nobody mourns their loss.

Questions

A

1. Express the following centigrade temperatures in degrees Fahrenheit: 100, 50, -40 , 0, 15.
2. Express the following Fahrenheit temperatures in degrees centigrade: 68, 212, 113, -40 , 95.
3. Construct a graph from which, for any temperature between 0° C. and 100° C., you can read the corresponding Fahrenheit temperature.

B

1. The distance between the upper and lower fixed points on a thermometer is 15 cm. At what height above the lower point would the mercury stand when the temperature is 20° C.?
2. You are provided with a mercury thermometer with an ungraduated scale. Describe how you would use it to find the temperature of the room. *Lond.**
3. Why would a thermometer constructed similarly to an ordinary laboratory thermometer be useless for taking a patient's temperature? *O.C.**
4. Why is mercury a suitable liquid to use in a thermometer? Why is alcohol to be preferred? Why is water unsuitable? Describe a thermometer which registers both maximum and minimum temperatures and explain its mode of action. *Lond.*

CHAPTER 14

EXPANSION OF SOLIDS

Recapitulation. When studying the expansion of solids in our earlier book we saw first of all that a solid when heated expands only very slightly, and that rather special apparatus is needed to detect it. Figs. 14/1*a*, 1*b* and 14/2 will recall some of the apparatus we used.

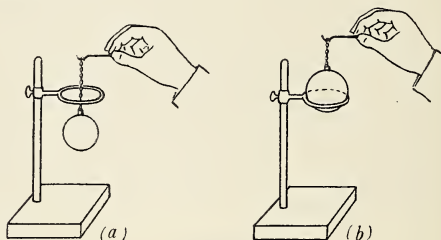


FIG. 14/1

But though the amount of expansion is very small, we saw that it was capable of being put to certain practical use. A steel tyre can be made to fit a cart wheel very tightly by making it of such a size that it can only be put on when it is

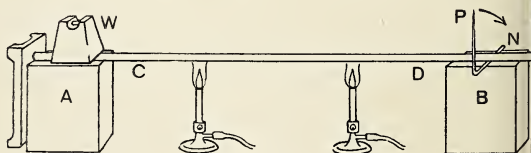


FIG. 14/2

hot. On cooling, it contracts and grips the wheel very tightly indeed. This principle of 'shrinking on' is used in fixing the wheel of a railway wagon on its axle. It is also used

setting boiler plates together, the rivets being put in when hot and then hammered.

A glass stopper which has become set fast in a bottle may be removed by gently heating the neck of the bottle so as to make it expand, and then tapping the stopper.

We also considered some cases in which expansion with change of temperature is a source of trouble. The engineer has to make allowance for it in building bridges, and we have met with cases in which thick glass tumblers, jam jars, etc., have been broken through pouring hot liquid into them. Generally speaking, the thicker the glass, the more easily it is broken by unequal heating. An extreme example is seen in the new reflecting telescope at Palomar (California). The great mirror, 200 in. in diameter, is 30 in. thick. It was made of Pyrex glass which has a very low coefficient of expansion (p. 139). Even so it was necessary, after casting the glass, to spread the cooling process over a period of *ten months*!

Further Applications. The rate of a pendulum clock is controlled by the pendulum, which swings more slowly if its length is increased. The length is measured from the point of suspension to the centre of gravity of the bob.

Thus to regulate a clock which is gaining, we lower the bob by the help of a little screw fixed just below it. If the clock is losing we raise the bob.

Now, if an ordinary pendulum clock has been keeping good time in the winter, it will begin to lose as the days become warmer. The reason is obvious—the length of the pendulum has increased.

There are various ways of avoiding the trouble. In some cases the pendulum bob is made to consist of a glass cylinder (two cylinders side by side) filled with mercury. Suppose (Fig. 14/3) S is the point of suspension of the pendulum and G the c.g. of the mercury. If the pendulum is always to swing at a constant rate, the distance SG must remain constant. Now with increase of temperature the rod SB becomes longer, and this would cause SG to increase. But the same increase of temperature causes the mercury to expand, and



FIG. 14/3

therefore its c.g. G is raised, causing SG to decrease. By properly adjusting the quantity of mercury the two opposing tendencies may be made to cancel out, and the clock keep excellent time.

(Q. *Would it be possible for such a clock to gain with increase of temperature and vice versa? If so, how could it be put right?*)

Alloys such as 'invar' are now made which undergo practically no change of length¹ as the temperature varies, and if these are used there is no need for the rather costly cylinder of mercury.

The rate of a watch is controlled by a balance wheel

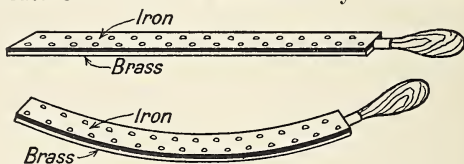


FIG. 14/4

Expansion obviously causes the weight of the rim to be moved farther from the centre, and this slows down the rate of movement.

To understand the principle of compensation, we must go back to the compound-bar experiment discussed in the early book. When such a bar is heated, it bends, with the more expansible brass on the outside of the bend (Fig. 14/4).

A 'compensated' balance wheel has its rim made of two metals, brass melted on to steel, the brass being outside. The wheel has two steel spokes OA and OB in line with one another (Fig. 14/5). With rise of temperature these expand, and this of itself would throw the weight of the rim farther out from the centre. But the rise of temperature also causes an increased amount of bending, and this

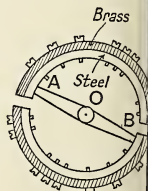


FIG. 14/5

¹ Its coefficient of expansion (see p. 139) is about $\frac{1}{11}$ th that of steel, and so for a very high degree of accuracy a little compensation is still required. This is secured by so arranging a small brass cylinder that as it expands, it raises the bob.

ngs the weight of the rim nearer to the centre. By correctly adjusting the weight of the rim (which is done by attaching screws to it), the two effects can be made to cancel out.

Q. If a watch with such a balance wheel were found to gain in hot weather, can you suggest what might be done?

Coefficient of Expansion. So far we have been content with such statements as that 'brass when heated expands more than iron.' We can easily see, however, that a *long* strip of iron would probably expand more than a *short* strip of brass. Even if we take brass and iron strips of equal length, it seems likely that an iron strip strongly heated would expand more than a brass strip only slightly heated. In fact, when we wish to speak with exactness about the expansion of a substance, we must have regard to (i) the length of the strip at which it is being heated, and (ii) the rise of temperature. If the length of the strip or wire is 1 centimetre and the rise of temperature is 1°C. , the increase in length is known as the *coefficient of expansion*.

Example 1. *If 1 centimetre of iron wire expands 0.000012 cm. for a rise of 1°C. of temperature, what would be the expansion (measured in feet) of 1 ft. of iron wire, for the same rise of temperature?*

Suppose 1 ft. = x cm.

1 cm. for a rise of 1°C. expands 0.000012 cm.

x cm. " " " 0.000012 x cm.

1 ft. " " " 0.000012 ft.

Evidently the number representing the coefficient of expansion will be the same whether we measure the original length in centimetres and also measure the expansion in centimetres, or measure the original length in feet and the expansion in feet; and similarly for any other units, such as inches or even miles. Accordingly, we define the **coefficient of expansion of a substance** as the increase which unit length of that substance undergoes when its temperature is raised 1° Centigrade.

It will be noticed that so far we have dealt only with measurements of *length*. The coefficient of expansion just defined is really the coefficient of *linear* expansion, and is so expressed. The question of the expansion of surfaces (e.g. of a sheet of copper foil) and of volume will be dealt

with presently. Evidently the term 'coefficient of linear expansion' can only be applied to a solid.

Let us now see how the coefficient of linear expansion of substance can be determined.

One method is illustrated in Fig. 14/6. The metal in the form of a tube is supported by a wooden frame, one end being fixed, while the other passes through a hole in the frame so that the tube is free to expand. Round this hole, fixed to the frame, is a metal plate in the face of which three small holes have been drilled to take the feet of a spherometer while the central screw of the latter touches a flat cap which closes the tube at Q. The tube is wrapped round with felt (not shown in the figure), and steam can be passed through it as indicated.

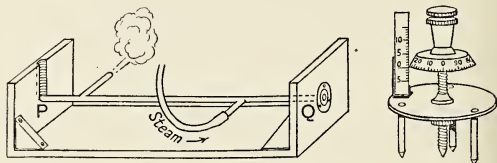


FIG. 14/6

To carry out an experiment, the spherometer is placed in position and adjusted till the central screw just touches the metal cap. The reading is then taken and the spherometer is removed. The temperature (room temperature) is recorded, after which steam is passed through the tube for a few minutes. While it is still passing, another spherometer reading is taken. To make sure that expansion is complete, steam is passed for another two minutes. There should of course, be no further change in the reading. The length of the tube is measured, and the result is then worked out as follows:—

Length of tube	80 cm.
Room temperature	15° C.
Spherometer reading (i) 2.76 mm. (ii) 4.04 mm.	
(iii) 4.04 mm.	

∴ Expansion = 1.28 mm. = 0.128 cm.

Rise of temp. = $100 - 15 = 85$ deg.

∴ expansion of 80 cm. for $85^\circ = 0.128$ cm.

∴ „ „ 1 cm. „ $1^\circ = \frac{0.128}{80 \times 85} = 0.0000188$,

coefficient of linear expansion = 0.0000188 .

Calculations. Let us now consider one or two examples relating to coefficient of expansion.

Example 2. *An engineer decides to allow for an extreme variation 70° C. in atmospheric temperature. How much allowance should he make for the expansion of the iron girders of the Forth bridge, 1710 feet long? (Coeff. of linear expansion of iron = 0.000012 .)*

1 foot of iron for a rise in temperature of 1° C. expands 0.000012 ft.

1710 feet for the same rise expand 0.000012×1710 ft.

1710 feet for a rise of 70° C. expand $0.000012 \times 1710 \times 70$
 $= 1.436$ ft. or **17.2 in.**

N.B. The bridge expands in sections, each of which would be its proper proportion of this total.

Example 4. *A surveyor's steel chain has a length of exactly 22 yards at 60° F. What will be its error if used for making measurements at 15° F.? (Coefficient of linear expansion of steel = 0.000012 per deg. C.)*

The difference in temperature is 45 deg. F.

Now 9 divisions on the Fahrenheit thermometer correspond to 5 divisions on the Centigrade

∴ 45 divisions on the one correspond to 25 on the other.

22 yd. of steel, for a fall of 1 deg. C., contracts 0.000012 yd.

∴ 22 yd. for a fall of 25 deg. C. contract

$0.000012 \times 22 \times 25 = 0.0726$ yd.

$= 2.61$ in.

N.B. It was the fact that surveyor's chains were found to vary appreciably in length with change of temperature, that led scientists in the 18th century to pay attention to the subject of coefficient of expansion.

Coefficient of cubical expansion. The expansion of unit volume of a substance for a rise in temperature of 1° C. gives the coefficient of *linear* expansion. Similarly the expansion of unit *volume* of a substance for this same rise of temperature gives us the coefficient of *cubical* expansion. *The coefficient of cubical expansion of a substance is the amount by which the volume of that substance expands for a rise in temperature of 1° Centigrade.*

There is a simple relation between the coefficient of cubical expansion of a substance and its coefficient of linear expansion. Let us see what it is. Suppose the coefficient of linear expansion of a certain substance is l , while its coefficient of cubical expansion is c .

Fig. 14/7 represents a cube of the substance in question and each edge of the cube measures 1 unit of length.

\therefore volume of cube = 1 unit of volume.

Let the temperature of the cube be raised by 1°C .

Then length of each edge increases from 1 unit to $(1 + l)$ units.

\therefore volume becomes $(1 + l)^3 = 1 + 3l + 3l^2 + l^3$.

But original volume of cube was 1 unit.

\therefore increase = $3l + 3l^2 + l^3$.

Now because l is a very small quantity, $3l^2$ may be neglected (e.g. for steel, $l = 0.000012$, and $3l^2 = 3 \times 0.000012^2 = 0.00000000432$), and l^3 is smaller still.

\therefore increase = $3l$ (practically).

But the increase in volume of this unit cube for a rise of temperature of 1°C . is our coefficient of cubical expansion c .

$$\therefore c = 3l$$

Thus the coefficient of cubical expansion of a substance is numerically equal to three times its coefficient of linear expansion.

You can easily work out for yourself the corresponding relation between the coefficient of surface expansion (s) of a substance and its coefficient of linear expansion l . You will find that, numerically, $s = 2l$.

Example 5. The temperature of a piece of copper weighing 895 grams is raised from 10°C . to 100°C . Find the increase in volume (Density of copper = 8.95 gm. per c.c. Coefficient of linear expansion = 0.000017.)

Weight of copper = 895 gm.

\therefore volume = $895 \div \text{density} = 895 \div 8.95 = 100 \text{ c.c.}$

Coefficient of cubical expansion = $3 \times 0.000017 = 0.000051$

i.e. 1 c.c. of copper for rise of 1 deg. expands 0.000051 c.c.

\therefore 100 c.c. for rise of 90° expands

$$0.000051 \times 100 \times 90 = 0.459$$

Why do substances expand on heating? Suppose a large floor is divided up into squares of, say, 2 ft. side. At each

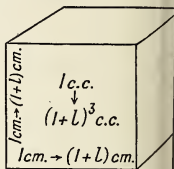


FIG. 14/7

int where lines intersect—points such as A, B, C, D, E, F—person is standing, and all these people are engaged in some regular side-to-side exercises. Their arms and legs are moving inwards and outwards, but not very far, because of contact with neighbouring persons.

Now, suppose the order is given that this side-to-side movement is to be increased. This requires more room, and the people must open out. As a result, the total size of the crowd will be greater than it was before.

This illustration will help us to understand what happens when a solid substance is heated. The extremely small bits (molecules) of which the solid is composed have a certain regular arrangement, represented roughly by the people standing at the corners of the squares. The molecules are vibrating, and this is represented by the side-to-side movement of the people.

When the substance is heated, the side-to-side movement of the molecules is increased, and they need more room. The molecules 'open out,' and the result is expansion.

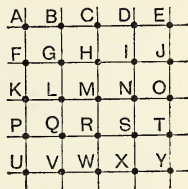


FIG. 14/8

Questions

Explain the term *coefficient of linear expansion*, and describe an experiment to find its value for brass.

A metal ball has a diameter of 4.02 cm., and a hole in a brass plate has a diameter of 4.00 cm. when both are at 20° C. To what temperature must the plate be heated so that the ball (at 20° C.) may just pass through the hole? (The coefficient of linear expansion of brass is 0.000018 per deg. C.) *Lond.*

If the coefficient of linear expansion of steel is 0.000012 per Centigrade degree, what will be its value for a Fahrenheit degree? If the length of a steel rail when laid at 56° F. is 30 ft., what interval must be left between the rails to allow for a rise in temperature to 120° F.? *Camb.**

To what temperature must two rods of copper and iron respectively, each 100 cm. long at 15° C., be heated in order that they may differ in length by 1 mm.? (The coefficients of linear expansion of copper and iron are 0.000018 and 0.000012 respectively, both per deg. C.) *Lond.**

4. Define *coefficient of linear expansion*. Give two examples of practical importance in which use is made of the linear expansion of metals with temperature rise.

The length of a mercury column as read on a brass scale is 76 cm., the temperature being 20°C . If the scale was graduated when the temperature was 0°C ., what is the correct length of the column? (Coefficient of linear expansion of brass is 0.000017 per deg. C.) *Brist.*

5. If a centimetre cube of copper at 0°C . has a mass of 8.9 gm., find its volume and density at 100°C ., given that the coefficient of linear expansion of copper = 0.000017 per deg. C. *Camb.**

6. An iron pipe is 10 ft. long at 60°F . Calculate its length when steam at 300°F . passes through it. (Coefficient of linear expansion of iron = 0.000012 per deg. C.). *C.W.B.**

7. Two bars A and B of different metals lie parallel, side by side, with two ends adjacent to each other and held rigidly there. The other ends are free. The longer bar A is 150 cm. long and the coefficients of linear expansion of the bars A and B are 0.000012 and 0.000018 respectively. What must be the length of B so that the difference in length of the bars is always the same irrespective of temperature? *Dur.**

8. The internal volume of a sealed hollow brass cube at 0°C is 1,000 c.c. and the coefficient of linear expansion of brass is 0.000018 per deg. C. Find (a) the volume of the cube when it is placed in an oven at 200°C ., (b) the density of the gas inside it at this temperature if the density in the cube at 0°C . is 3.5 gm. per litre. *J.M.B.**

CHAPTER 15

EXPANSION OF LIQUIDS

the last chapter we saw how to measure the coefficient of linear expansion of a solid. We have then only to multiply the result by 3 to obtain the coefficient of cubical expansion.

There is no such thing, however, as the coefficient of linear expansion of a liquid. We might put the liquid in a long tube (Fig. 15/1) and measure the length of the column AB at a certain temperature. We could then heat it to another known temperature and measure the amount of expansion.

It looks as though we might now work out the coefficient of linear expansion in much the same way as, on p. 40, we worked out the coefficient of linear expansion of a metal. Why not?

If AB were a bar of copper or other solid substance and we heat it, it expands both lengthways and sideways, and we can find its coefficient of linear expansion because its expansion sideways in no way affects its expansion lengthways.

The sideways expansion of the liquid, however, is resisted by the walls of the containing vessel, and therefore, shows itself as lengthways expansion. In Fig. 15/1 BC is a measure of the *total* expansion of the liquid in all directions, i.e. it is a measure of the *cubical* expansion, not of the linear.



FIG. 15/1

The last statement requires further qualification.

We cannot heat the liquid without also heating the tube containing it. This causes the tube to expand, and so the apparent expansion BC is less than the real expansion. If BC were 10 c.c. and the glass tube itself has expanded 1 c.c., it is evident that the *real* expansion is 10 c.c. In fact, we cannot find the amount by which the liquid has really expanded without knowing that by which the containing vessel has expanded.

However, we will leave this point for the moment, and try to find the 'coefficient of apparent expansion' of a liquid in a glass vessel. Suppose we use glycerine.

Take a density bottle, see that it is dry, and then weigh (i) empty, (ii) full of glycerine. The bottle and glycerine should be at room temperature, so when filling take care to hold the bottle only by the neck. There should be no air bubbles.

By means of thread, suspend the bottle in a beaker of water as shown in Fig. 15/2. This water should be at room temperature, so it will be better if the beaker has been filled some time before the lesson.

Take the temperature, heat the water to boiling point and let it continue boiling gently for five minutes. Lift out the density bottle, let it cool, wipe it dry and weigh it. The results are then worked out as follows. Suppose

Wt. of bottle	15.26 gm.
Bottle + glycerine before heating	78.37
" " after heating	75.80
Initial temperature	15° C.
Final temperature (assumed)	100° C.
Wt. of glycerine left in bottle	$75.80 - 15.26 = 60.54$ gm.
Wt. that has escaped by expansion	$78.37 - 75.80 = 2.57$ gm.

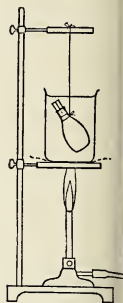


FIG. 15/2

Let s = sp. gr. of glycerine at 15° C.

Then *volume* of liquid left in bottle, if allowed to cool to 15° would be $\frac{60.54}{s}$ c.c.

But at 15° C. wt. of glycerine required to fill bottle
 $= 78.37 - 15.26 = 63.11$

\therefore volume of bottle $= \frac{63.11}{s}$ c.c.

\therefore volume of space A (Fig. 15/3) $= \frac{63.11}{s} - \frac{60.54}{s} = \frac{2.57}{s}$

Now the glycerine left in the bottle after the experiment ($60.54/s$ c.c.) if heated once more to 100° C. would evidently expand just sufficiently to fill the bottle.

i.e. $\frac{60.54}{s}$ c.c. for rise of 85 deg. expands $\frac{2.57}{s}$ c.c.

\therefore 1 " " " " " $\frac{2.57}{s} \div \frac{60.54}{s} = \frac{2.57}{60.54}$ c.

\therefore 1 " " 1 " " $\frac{2.57}{60.54 \times 85} = 0.000496$

i.e. mean coefficient of apparent expansion of glycerine in g between 15° and 100° = **0.000496**.

N.B. (i) 'mean' coefficient, because we do not know that it is expanded *equally* for each degree of increase in temperature. We have simply taken the total expansion for 85 degrees, and divided by 85.

(ii) If asked to find the mean coefficient of expansion say between 30°C. and 60°C. , we should (before taking our first weighing of bottle + glycerine) have to leave the bottle surrounded for some time with water at 30°C. , then wipe it and weigh it. Afterwards, of course, we should bring the temperature to 60°C. instead of 100°C. , as in the experiment discussed above.

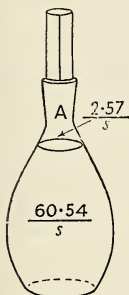


FIG. 15/3

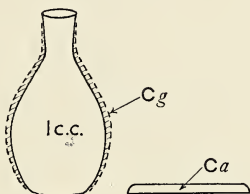


FIG. 15/4

The relation $c_r = c_a + c_g$. Imagine a bottle whose capacity is exactly 1 c.c., filled with a certain liquid. The temperature is raised by 1°C. , and a volume of liquid will be expelled equal to c_a , where c_a is the coefficient of apparent expansion of the liquid.

If the coefficient of expansion of the glass is c_g , the bottle itself has expanded by a volume c_g for this 1° rise of temperature. Therefore the liquid must have expanded by (the amount expelled) *plus* c_g (the amount required to fill the extra space in the bottle), i.e. the coefficient of *real absolute* expansion of the liquid, c_r , is equal to $c_a + c_g$. We may therefore write $c_r = c_a + c_g$.

Fig. 15/4 may help to give a clearer idea of the argument. The 1-c.c. bottle is exaggerated. Its expansion c_g is represented by the shaded portion, and the expelled liquid of volume c_a is represented as piled up on the table.

How c_g is found. We have seen how c_a is found for glycerine. In a similar way it could be found for *mercury*.

Now, for *mercury*, the coefficient of absolute expansion c_r , has been found with great exactness to be 0.00018. Thus if we want to find c_g for the glass of a given density bottle, we use the latter to find c_a for mercury by the method already described for glycerine. Suppose we find that c_a is 0.000155. Then from $c_r = c_a + c_g$ we have $c_g = c_r - c_a = 0.000181 - 0.000155 = 0.000026$.

We can now find c_r for glycerine. For we have already found c_a , we know c_g , and $c_r = c_a + c_g$.

So long as we keep to the same density bottle, for which c_g is known, we can of course find c_r for any other liquid, first finding c_a as already described, and then making use of the relation $c_r = c_a + c_g$.

Example. If the real coefficient of expansion of a certain liquid is 0.00138 per deg. Centigrade, find the value per degree Fahrenheit.

A flask is exactly filled by 1000 c.c. of the liquid at 0°C . The coefficient of linear expansion of the glass is 0.000004. What volume will overflow if it is heated to 60°C ?

When 1 c.c. of the liquid has its temperature raised 1°C , it expands by 0.00138 c.c.

For a rise of 1°F , the expansion would be only five-ninths of that, i.e. 0.00078.

The coefficient of cubical expansion of the glass is 3 times the linear = $3 \times 0.000004 = 0.000012$.

The amount of liquid overflowing evidently depends on the expansion relative to the glass, i.e. on its coefficient of apparent expansion. This will be $c_r - c_g = 0.00138 - 0.000012 = 0.001368$.

Thus, relative to the glass,

1 c.c. for 1 deg. expands 0.001368 c.c.

\therefore 1000 „ 60 „ „ 0.001368 \times 1000 \times 60
= 82.08 c.c.

i.e. the amount overflowing will be 82.08 c.c.

¹ A very broad indication of the method may be given. It depends on the U-tube method for finding the density of a liquid (p. 93), instead of having a column of water balanced by (say) one of alcohol, we have a column of mercury at 0°C . balanced by another column of mercury at some definite temperature $t^\circ \text{C}$. The two densities are now to be compared, for $d_0/d_t = h_t/h_0$, where h_t and h_0 are the heights of the mercury columns at t° and 0° respectively. Thus d_0/d_t is found, and it can easily be shown that $d_0/d_t = 1 + c_r t$. Hence, as we know d_0 , we can calculate c_r .

Applications. The thermometer provides us with one of the most important applications of the expansion of a liquid. Another interesting application is seen in one type of *thermostat*. In Fig. 15/5 a bath of liquid L is being heated, and it is desired to keep it at a constant temperature. To secure this, the lower part of a thermostat is immersed in the liquid. The thermostat contains toluene, a liquid with a very high coefficient of expansion.

Notice that the burner is supplied with gas by two paths: (i) the path ACBDE, and (ii) the by-pass. The adjustment is such that when the liquid in the bath reaches the desired temperature, the mercury, pushed by the expanded toluene, just covers the tip B, so closing the path ACBDE. Enough gas, however, goes through the by-pass to keep the burner alight. When the temperature begins to fall, the toluene contracts, the mercury falls, tip B is uncovered, and the gas supply is reinforced.

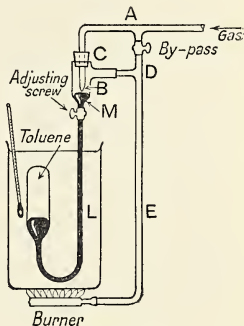


FIG. 15/5

Questions

1. When a lake freezes, why does ice first form on the surface, while the temperature some distance below the ice is higher than the freezing point? *Brist.**
2. A mercury thermometer with its fixed-point markings separated by a distance of 25 cm. has a bulb with a volume of 10 c.c. up to the 0°C . mark. What is the mean diameter of the stem of the thermometer? The coefficient of apparent expansion of mercury in glass = 0.00015 per deg. C. *Lond.**
3. A specific-gravity bottle when full contains 27.5 gm. of cerine at 16°C . What weight of glycerine will be expelled when the bottle and contents are heated to 90°C .?
Coefficient of real expansion of glycerine = 0.00053 per $^{\circ}\text{C}$.
Coefficient of linear expansion of glass = 0.0000090 per $^{\circ}\text{C}$.
*C.W.B.**
4. A glass flask weighs 80 gm. empty and 580 gm. when full of water at 20°C . What does it weigh when full of water at 100°C .? (The apparent coefficient of expansion of water = 0.0004 per $^{\circ}\text{C}$.) *O.C.**

5. The bulb and stem of a mercury thermometer contain 0.6 c.c. of mercury when the temperature is 0°C . What is the distance between the 0°C . and the 100°C . marks if the internal diameter of the stem is 0.2 mm.? (Coefficient of apparent expansion of mercury is 0.000156 per $^{\circ}\text{C}$.) *Brist.**

6. Distinguish between the coefficients of *real* and *apparent* expansion of a liquid. What is the relation between them?

In an experiment to measure one of these coefficients, two columns of the liquid are balanced in vertical glass tubes connected by a horizontal tube at their lower ends. The temperature of one column is 0°C . and that of the other 100°C . The heights of the columns are 50 cm. and 54.5 cm. Which coefficient is measured by this method? Evaluate it, giving the theory and your method of calculation. *Dur.*

7. The cylindrical bulb of a mercury thermometer is 2 cm. long and has an internal diameter of 5 mm. The diameter of the stem is 0.2 mm. If the bulb is full at 0°C ., what must be the minimum length of the stem to measure temperatures up to 110°C .? (The coefficient of real expansion of mercury is 0.000184 per $^{\circ}\text{C}$. and the coefficient of linear expansion of glass is 0.0000080 per $^{\circ}\text{C}$.) *Lond.**

8. A graduated tube contains mercury which at 0°C . occupies 100 divisions. What will be the reading if the whole is raised to 100°C .? (Coefficient of absolute expansion of mercury = 1.8×10^{-4} , coefficient of linear expansion of glass = 8×10^{-6} per $^{\circ}\text{C}$.) *O.C.**

CHAPTER 16

EXPANSION OF GASES

We already know that air expands when it is heated. A flattened ping-pong ball, for instance, can be straightened out by warming it in front of the fire. Other gases too expand when they are heated, and two questions at once suggest themselves.

1) Exactly how much does air expand when heated? In other words, what is its coefficient of expansion?

2) Have different gases different coefficients of expansion (we have found to be the case with solids and liquids) or are they all alike in this respect?

Measurement of Expansion. First we had better be quite clear about the meaning of *coefficient of expansion of air* ('coefficient of expansion of air at constant pressure'). It is the *expansion for a rise in temperature of 1° C. of a quantity of air whose volume at 0° C. is 1 c.c., the pressure remaining constant.*

The last four words are very important. Moderate or even great changes of pressure make no appreciable difference in the volume of a solid or liquid, but we know from common experience, or from Boyle's law, that this is far from being the case with a gas.

Note that the definition does not imply that the change in temperature is necessarily from 0° C. to 1° C. It could be from 17° to 18° , or 99° to 100° , or -25° to -24° . The point is that the quantity of air must be such that its volume at 0° C. would be 1 c.c.

Various methods are in use for finding this coefficient. The one to be described we shall use what is known as *Gay's Apparatus*. Here the air under examination is confined in the bulb A and in the upper part of the tube BC, which is graduated so that the exact volume can be read off. This is connected with the tube DE, open to the atmosphere, and also with the tube FGH, which is closed with the tap H. The apparatus is filled with sulphuric acid to about B. The acid serves to keep the imprisoned air quite dry.

The apparatus is placed in a beaker of suitable size, and to begin with may be surrounded by water containing quantity of ice to bring the temperature to 0°C . Acid is now run out at H until the acid stands at the same level in DE and BC. As DE is open to the atmosphere, the air enclosed in A will now be at atmospheric pressure. The volume is noted, and the temperature (0°C).

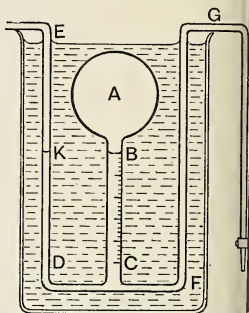


FIG. 16/1

The beaker is now heated, the water being gently stirred, till the temperature has risen 15 or 20 deg. The acid surface in the bulb tube will now be lower than in the open one, so acid is once more run out at H until the two are at the same level. The volume and temperature are again noted. Proceeding in this way we obtain a series of readings until, say, 80° or even 100° is reached (this last reading presents certain difficulties in practice), the pressure in every case being adjusted to atmospheric. Suppose the following results have been obtained:—

Volume (c.c.)	.	.	35.2	37.8	40.3	42.9	45.5	48.1
Temp. ($^{\circ}\text{C}$.)	.	.	0	20	40	60	80	100

Let us work out the mean coefficient of expansion between 0° and 100° .

35.2 c.c. for a rise of 100° expands $(48.1 - 35.2) = 12.9$ c.c.

$$\therefore \quad 1 \text{ c.c.} \quad \text{,,} \quad \text{,,} \quad 1^{\circ} \quad \text{,,} \quad \frac{12.9}{35.2 \times 100} \\ = 0.00366 \text{ or } 1/273$$

Thus the mean coefficient of expansion between the temperatures is $1/273$. The results are almost exactly the same between 0° and 40° , 0° and 60° , etc.

In short, our experiment indicates that *if air is heated constant pressure, then for each increase of temperature of $C.$ it expands by $1/273$ rd of its volume at $0^{\circ} C.$*

We may note here that if the experiment is carried out with other gases, the coefficient of expansion is still found to be $1/273$. Thus while different solids and different liquids have their own special coefficients of expansion, all gases have the same—namely, $1/273$.

Let us next express our results graphically.

Our $100^{\circ} C.$ reading (volume 48.1 c.c.) is represented by the point P, and our $0^{\circ} C.$ reading (volume 35.2 c.c.) by the

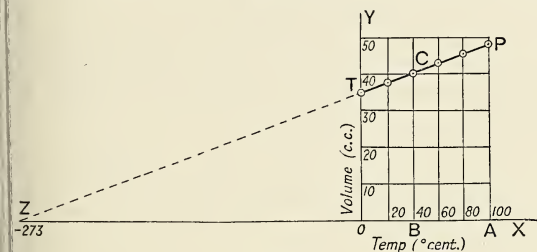


FIG. 16/2

at T. Between these points we notice that the volume increases with almost perfect regularity (the slight deviations actually due to errors of experiment). The dotted line TZ shows what would happen *if the volume continued to decrease at the same rate*. At a temperature of $-273^{\circ} C.$ the volume would become nothing at all!

We feel that this is a rather impossible state of things, and need it does not occur. What *does* happen is that at about $-90^{\circ} C.$ the air liquefies,¹ and with further fall of temperature the contraction proceeds at a much slower rate than when we were dealing with the air as a gas.

$-273^{\circ} C.$, then, is the temperature at which the volume of a gas would become zero if it continued to contract at the

For some time before this temperature is reached, the air is found to be contracting to a greater extent than would be expected from the

same rate as it does when well above its temperature liquefaction. For reasons which will be mentioned later, the temperature -273°C. is known as *absolute zero*. 'Absolute temperature' means the number of degrees (Centigrade) above absolute zero. For instance, 0°C. is 273°C. above absolute zero, and so $0^{\circ}\text{C.} = 273^{\circ}\text{Absolute temperature}$. Similarly $-200^{\circ}\text{C.} = 73^{\circ}\text{Absolute}$, because it is 73° above absolute zero, $100^{\circ}\text{C.} = 373^{\circ}\text{Absolute}$, and so on. In fact, *to bring degrees centigrade to degrees absolute, add 273*. Not a very hard rule!

Notice that on our graph absolute temperatures are represented by horizontal distances from Z. OA, for instance, represents the centigrade temperature 100° , but $ZA = ZO + OA = 273 + 100$, so ZA represents the corresponding *absolute* temperature. Similarly, OB represents the centigrade temperature 40° , but $ZB = ZO + OB = 273 + 40$, which is the corresponding absolute temperature.

Now, $\frac{AP}{ZA} = \frac{BC}{ZB}$, either from similar triangles, or from trigonometry (each ratio being equal to $\tan Z$).

$$\text{Hence } \frac{AP}{BC} = \frac{ZA}{ZB}.$$

But AP and BC represent two volumes, while ZA and ZB represent the corresponding absolute temperatures. Hence *the volumes are proportional to the absolute temperatures*.

ZA and ZB of course represent two *particular* absolute temperatures (373 and 313 respectively), but the reason obviously applies to *any* two.

It has been understood throughout that the volumes measured at constant pressure. Keeping this in mind, we may now re-state **Charles' Law** in the following form, which should be carefully committed to memory. *If the pressure of a given mass of gas remains constant, its volume is proportional to its absolute temperature.*

In mathematical form, we may express this by the equation

$$\frac{V_1}{V_2} = \frac{T_1}{T_2}, \text{ where } V_1 \text{ is the volume of the gas at absolute temperature } T_1, \text{ and } V_2 \text{ its volume at absolute temperature } T_2.$$

Making use of the idea of absolute temperatures, calculations on volume changes in gases are very easy. For instance:

Example 1. *An enclosed quantity of gas occupies 27 c.c. at a temperature of 15° C. What would be its volume at 87° C., the pressure remaining unaltered?*

$$15^{\circ} \text{ C.} = (273 + 15)^{\circ} \text{ Abs.} = 288^{\circ} \text{ Abs.}$$

$$87^{\circ} \text{ C.} = (273 + 87)^{\circ} \text{ Abs.} = 360^{\circ} \text{ Abs.}$$

Thus in our equation, $V_1 = 27$, $V_2 = ?$ $T_1 = 288$, $T_2 = 360$.

$$\therefore \frac{V_1}{V_2} = \frac{T_1}{T_2} \text{ becomes } \frac{27}{V_2} = \frac{288}{360}$$

$$\therefore V_2 = \frac{27 \times 360}{288} = 35 \text{ c.c.}$$

When the pressure also varies. We have just seen that Charles' law tells us how the volume of a gas varies with the temperature, the pressure remaining constant, and that it

may be expressed as $\frac{V_1}{V_2} = \frac{T_1}{T_2}$. Now Boyle's Law (p. 101)

tells us how the volume of a gas varies with the pressure, the temperature remaining constant, and is expressed mathematically as $P_1V_1 = P_2V_2$. Both laws are expressed by the

equation $\frac{P_1V_1}{T_1} = \frac{P_2V_2}{T_2}$. For if the pressure is constant,

$P_1 = P_2$ and the equation reduces to $\frac{V_1}{T_1} = \frac{V_2}{T_2}$ or $\frac{V_1}{V_2} = \frac{T_1}{T_2}$, i.e.

Charles' Law. If the temperature is constant, we have $T_1 = T_2$ and the equation becomes $P_1V_1 = P_2V_2$, i.e. Boyle's

Law. In fact, the equation $\frac{P_1V_1}{T_1} = \frac{P_2V_2}{T_2}$ covers all cases.

Let us use it to solve a problem in which there is a variation of pressure and temperature.

Example 2. *A litre of air at N.T.P.¹ weighs 1.29 gm. What would it weigh at 91° C. and 38 cm. pressure?*

Let us first find what volume the air would occupy under the given conditions.

I.e. normal temperature and pressure (0° C. and 76 cm. pressure). Sometimes the letters S.T.P. (Standard Temperature and Pressure) are used instead.

$$\begin{aligned} V_1 &= 1 \text{ litre, } P_1 = 76 \text{ cm., } T_1 (0^\circ \text{ C.}) = 273^\circ \text{ Abs.} \\ V_2 &= ? \quad P_2 = 38 \text{ cm., } T_2 (91^\circ \text{ C.}) = 91 + 273 = 364^\circ \text{ Abs.} \end{aligned}$$

$$\begin{aligned} \frac{P_1 V_1}{T_1} &= \frac{P_2 V_2}{T_2} \\ \therefore \frac{76 \times 1}{273} &= \frac{38 \times V_2}{364} \\ \therefore V_2 &= \frac{76 \times 364}{38 \times 273} = 2\frac{2}{3} \text{ litres.} \end{aligned}$$

The weight, of course, has not changed,
i.e. $2\frac{2}{3}$ litres now weigh 1.29 gm.

$$\therefore 1 \text{ litre now weighs } \frac{1.29}{2\frac{2}{3}} = 0.484 \text{ gm.}$$

We obtain a somewhat shorter solution if we realise that the density of a gas is (i) directly proportional to the pressure, and (ii) inversely proportional to the absolute temperature.

$$\therefore \text{wt. of 1 litre} = 1.29 \times \frac{38}{76} \times \frac{273}{364} = 0.484 \text{ gm.}$$

Coefficient of increase of pressure. If in Lay's apparatus (p. 151) the volume of the enclosed air were 273 c.c. (actually it is only about one-tenth of this amount), and we raised the temperature from 0° C. to 1° C. , keeping the pressure constant, the air would expand by $1/273$ rd of 273, i.e. by 1 c.c., and the volume would therefore become 274 c.c.

Now, keeping the temperature at 1° , let us so increase the pressure that the volume is reduced to the original 273 c.c. If the original pressure is p_1 , and the new pressure required is p_2 , we have, by applying Boyle's Law, $274 \times p_1 = 273 \times p_2$

$$\therefore p_2 = \frac{274}{273} p_1 = p_1 + \frac{1}{273} p_1$$

i.e. the pressure would have to be increased by $1/273$ rd of its original value.

The net result is that the enclosed air has been heated from 0° to 1° at constant volume, and to secure this the pressure has had to be increased by $1/273$ rd of its value at 0° .

We have taken the simple case of 273 c.c. of gas heated from 0° C. to 1° C. , but by applying Boyle's Law in a similar way it can be shown generally that if an enclosed quantity of gas is heated from 0° C. to 1° C. , the pressure must be increased by $1/273$ rd of its value at 0° .

gas is heated at constant volume, then for each degree rise of temperature its pressure will increase by $1/273$ rd of the value it has at 0°C . This increase of pressure at constant volume, per degree rise of temperature, expressed as a fraction of the pressure at 0°C ., is known as the **coefficient of increase of pressure at constant volume**.

The fact that this coefficient must be equal to $1/273$ has so often been deduced from our knowledge that the coefficient of expansion at constant volume is also $1/273$. But we can obtain the result by direct experiment as follows.

The bulb A, 2 or 3 in. in diameter, is connected to the bent tubing BMC of very fine bore. At C this is connected to 5 or 6 ft. of pressure tubing CDE, terminating in the mercury reservoir F. This can be clamped in any desired position against a vertical support, and the height of the mercury surface can be read by means of the scale GH.

When the gas is surrounded by a vessel of crushed ice and after a suitable interval the reservoir is raised or lowered till the mercury stands exactly at a mark M scratched on the end of the narrow tube. The pressure must now be read. This is the pressure of the atmosphere (as recorded by the barometer) *plus* the amount by which F is higher than M. If F is below M the difference between the F and M levels must, of course, be *subtracted* from the barometer reading.

The gas is now surrounded by boiling water.

This causes the mercury to be pushed down to M, but by raising the reservoir it is forced back to that point by the increased pressure. The pressure is read as already described. Consider the following results:—

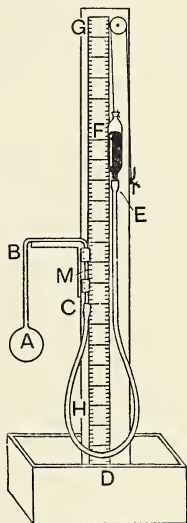


FIG. 16/3

Temp.	Position of F	Barometer reading	Pressure
0°C .	7.2 cm. below M	76.3 cm.	$76.3 - 7.2 = 69.1$
100°C .	18.1 cm. above M	76.3 cm.	$76.3 + 18.1 = 94.4$

The increase in pressure as the temperature is raised from to 100° is $94.4 - 69.1 = 25.3$ cm.

\therefore average increase per degree $= 25.3 \div 100 = 0.253$.

Expressed as a fraction of the pressure at 0° C. this is

$$\frac{0.253}{69.1} = \frac{1}{273}$$

i.e. mean coefficient of increase of pressure at constant volume between 0° C. and 100° C., is $1/273$.

After taking the reading at 100° C. we could allow the surrounding water to cool down, and take a series of temperature readings and the corresponding pressures, at intervals of 10 deg. or Working out the results we should obtain a series of values the coefficient of increase of pressure at constant volume, and with careful work these would be found to be remarkably constant at $1/273$.

The Constant-Volume Air Thermometer. We have just seen that if the temperature of the bulb of air is known, then reading the pressure we can calculate the value of the coefficient. Conversely, if we assume the value of the coefficient to be $1/273$, we can read the pressure and calculate the temperature. In short, the instrument can be used as a constant-volume air thermometer.

Suppose, for instance, we wish to find the temperature of a quantity of solid carbon dioxide. We put it in a suitable vessel, crush it, and surround A with it. We adjust the reservoir so that the mercury on the other side comes exactly to M, and take the reading. Suppose F is 27.1 cm. below the surface of the mercury, and that the barometer reads 76.3 cm. The pressure in the bulb is therefore $76.3 - 27.1 = 49.2$ cm. This is less than the pressure at 0° C. (69.1 cm.) by $69.1 - 49.2 = 19.9$ cm. Now for each fall of 1° C. the pressure falls by $1/273$ of the pressure at 0° C., i.e. by $\frac{1}{273}$ of $69.1 = \frac{69.1}{273}$ cm.

But the pressure has actually fallen by 19.9 cm.

\therefore no. of degrees below 0° C. is $19.9 \div \frac{69.1}{273} = 78.6$

i.e. the required temperature is -78.6° C.

We might work this out more easily by applying the equation $\frac{P_1 V_1}{T_1} = \frac{P_2 V_2}{T_2}$.

The 'V's' cancel, so $\frac{P_1}{T_1} = \frac{P_2}{T_2}$.

Now $P_1 = 69.1$ cm., $T_1 = 0^\circ \text{C.} = 273^\circ \text{A.}$, $P_2 = 49.2$ cm.

$$\therefore \frac{69.1}{273} = \frac{49.2}{T_2}$$

This gives $T_2 = \frac{273 \times 49.2}{69.1} = 194.4^\circ \text{Abs.}$
 $= (194.4 - 273)^\circ \text{C.}$
 $= -78.6^\circ \text{C. as before.}$

Advantages and disadvantages. One great advantage of the thermometer is that it can be used for recording extremely low temperatures, say down to -100°C. (we could use a thermometer containing hydrogen or helium instead of air or even lower ranges). Equally, it can be used for very high temperatures, especially if we use a bulb made of quartz; for these higher temperatures nitrogen is preferred. Another advantage is its accuracy. Temperatures obtained with the air thermometer agree exactly with those of an ideal gas thermometer scale devised by Lord Kelvin—the 'absolute'—to which reference will be made presently.

Its chief disadvantages as compared with the ordinary mercury thermometer are (1) greatly increased cost, (2) the process of finding a temperature is much more laborious, (3) we cannot easily use it to find the temperature of a *small* quantity of a substance.

The thermometer scale. You will remember that in making a centigrade mercury thermometer, we mark the 'lower fixed point' by the help of melting ice and call it 0°C. , and the 'higher fixed point' by the use of steam, calling this 100°C. Then we divide the space between the two marks into 100 equal parts.

Suppose we repeated the work, using sulphuric acid for thermometer liquid instead of mercury. On comparing the two thermometers we should find a very curious thing. At 0° and 100° they exactly agree, of course, but at intermediate

points they differ quite considerably. For instance, when the mercury thermometer records 50° the sulphuric acid thermometer records only 46° . We might say, 'Well, we are now exactly half-way between freezing point and boiling point. The mercury has moved up just fifty divisions, showing that it expands regularly. The acid has moved up only forty-six divisions, showing that it does not expand so much when heated from 0° to 50° as when heated from 50° to 100° .'

That is all very well, but—we are taking for granted that the mercury thermometer is correct. For anything we know, the sulphuric-acid thermometer may be correct, while the mercury is expanding more than it should at lower temperatures, and less than it should at higher ones. What are we to decide?

Lord Kelvin in 1848 devised a thermometer scale on an entirely new principle. 0° C. and 100° C. were defined in the usual way with reference to melting ice and boiling water, but he showed how to calculate 1/100th part of this interval, i.e. how to obtain a true interval of 1° C.—*without the aid of any thermometric substance whatever*. Because his scale is independent of any thermometric substance, it was called the **absolute** scale (Lat. *absolutum* = loosed from). On this scale, Kelvin showed that at -273° C. a body would be completely devoid of heat. We have met with this temperature before in connection with the contraction of gases. In fact, it is found that temperatures as recorded by the mercury thermometer agree almost perfectly with those of the absolute scale. Mercury also comes very well out of the comparison. Many other substances, however, including sulphuric acid, are found to expand by unequal amounts for equal increments of temperature.

Absolute zero. We have seen that from two points of view the temperature -273° C. attracts special attention. First, experiments such as that described on p. 152 indicate that at this temperature the volume of a gas would become zero if it continued to contract at the usual rate. Second, calculations by Kelvin show that at this temperature a substance would be completely devoid of heat.

Again, the electrical resistance of metals decreases as

perature is reduced, and in the neighbourhood of absolute—say at -272°C. —it practically vanishes. An electric current set up in a metal ring by means of magnetic induction continues to travel round and round the ring for days. Together, -273°C. is a very curious temperature.

Absolute zero has never been reached—probably never will be—but it has been very closely approached, usually in the course of experiments carried out to liquefy or solidify gases. In 1895 liquid air, boiling at about -190°C. , was obtained on a large scale by making use of the principle that when a compressed gas is allowed to expand there is a fall of temperature. In the same year Dewar succeeded in liquefying hydrogen by compressing it to 200 atmospheres, cooling it with the help of liquid air to -200°C. , and then allowing it to expand through a valve. The liquid hydrogen was at $-252\cdot78^{\circ}$. By rapid evaporation the temperature of the liquid was brought down to -259° , at which temperature it froze. With the help of liquid hydrogen, Kamerlingh Onnes in 1907 cooled compressed helium to -258° , and then allowed it to expand. Liquid helium was obtained by cooling at $-268\cdot82^{\circ}$, and by rapidly evaporating this under reduced pressure a temperature of $+0\cdot82^{\circ}\text{Abs.}$ was obtained. Finally, it was noted that certain substances become warmer when magnetised and cooler when demagnetised. By cooling them in boiling helium while in a magnetic field and then demagnetising, a temperature only $0\cdot0044^{\circ}$ above absolute zero was obtained.

Why is it impossible to have anything colder than -273°C. ? The molecules of which substances are composed are in constant movement. When a substance loses heat it loses energy—heat is a form of energy—and so the average energy of the molecules is reduced. Now, the energy of a molecule is represented by the product $\frac{1}{2}mv^2$, where m is its mass and v its velocity (p. 45), and when this energy is reduced there must fall in the value of v (m being constant). In the last case v becomes equal to zero. The molecule is at rest, and as there can be no further reduction of velocity there can be no lower temperature.

Questions

A

1. The following temperatures are given in degrees Centigrade 0° , 15° , -273° , 100° . What are the corresponding absolute temperatures?

2. The following temperatures are given in degrees absolute 0° , 100° , 273° . What are the corresponding centigrade temperatures?

3. A certain mass of gas, at 20° C., has a volume of 580 c.c. What would be its volume at (i) 0° C., (ii) 100° C.?

B

1. State Charles' law and describe an experiment to verify it. A thin glass bulb is sealed at 17° C., the air pressure inside being atmospheric. The maximum internal pressure that the bulb will withstand is 1.5 atmospheres. At what temperature will the bulb burst, the expansion of the glass being neglected. *Lon.*

2. If the density of nitrogen at 27° C. under a pressure of 76 cm. of mercury is 1.138 gm. per litre, what is its density at 100° C. the pressure remaining the same? *Dur.**

3. The gas required to fill a barrage balloon to a pressure of 75 cm. of mercury at 12° C. is stored in 10 cylinders each of volume 0.25 cu. metre. The gas in the cylinders is at a temperature of 15.8° C. and at a pressure of 84 standard atmospheres (one standard atmosphere = 76 cm. of mercury). Find the capacity of the balloon in cu. metres. *C.W.B.**

4. Describe how you would investigate the way in which the volume of a fixed mass of gas heated at constant pressure changes with the temperature.

The following results were obtained in a certain experiment:

Temp. ($^{\circ}$ C.)	.	10	40	60	85	95
Volume (c.c.)	.	26.0	29.0	31.0	33.5	36.0

Plot a graph of volume against temperature, arranging the axes so that temperatures between -300° C. and $+100^{\circ}$ C. are plotted along the X-axis, and volumes between 0 c.c. and 36 c.c. along the Y-axis. Produce the graph as far as you can beyond the plotted points. Comment briefly on the graph obtained.

5. On a day when the temperature is 10° C. a cycle tyre is pumped to a pressure of 80 cm. of mercury. What will be the pressure if the temperature rises to 35° C., the volume being assumed to remain constant? *Lond.**

7. Describe a constant-volume air thermometer, and explain how you would use it to find how the pressure of a mass of gas contained at constant volume varies with the temperature.

8. A high-altitude aircraft carries a supply of oxygen in cylinders at a pressure of 180 lb. per sq. in. measured at 27°C . What would be the pressure in the cylinders when the temperature inside the aircraft falls to -3°C ? *J.M.B.*

9. The pressure, expressed in cm. of mercury, of the gas in a constant volume air thermometer is 74 cm., 101.1 cm., and 94 cm. when the bulb of the thermometer is immersed in melting ice, boiling water and melting wax respectively. Find the temperature of the melting wax. *O.C.**

10. A sealed glass flask of air at atmospheric pressure and 20°C . is immersed in liquid air at -180°C . Neglecting any volume change in the flask, calculate the final pressure (in atmospheres) in the flask. *Lond.**

11. A mass of hydrogen occupies 200 c.c. at 17°C . and 780 mm. mercury pressure. What is its pressure when the volume is 100 c.c. and the temperature 27°C ? *Lond.**

12. A certain volume of oxygen occupies 570 c.c. at 12°C . and 57 cm. pressure. What would be its volume at N.T.P.?

13. 1 gm. of oxygen at N.T.P. has a volume of 700 c.c. What would be its volume at a temperature of 39°C . and a pressure of 10 cm.?

14. An enclosed quantity of air occupies 240 c.c. at 14°C . and 82 cm. pressure. What would be its volume at 28°C . and 10 cm. pressure?

15. The volume V of a certain mass of air was found to vary with the temperature t as follows:

$^{\circ}\text{C}$.	.	.	20.0	40.0	60.0	80.0	100.0
c.c.	.	.	97.7	104.6	111.0	117.1	124.3

Draw a graph of V against t , and deduce from it the volume of the gas at 0°C . *Camb.**

16. A vertical capillary tube, closed at its lower end, contains an air column of length 20.0 cm. with a mercury thread above it. When this tube is placed in melting ice, the air column contracts to a length of 18.9 cm.; when placed in steam, it expands to 22.1 cm. Find the laboratory temperature. *Camb.**

17. A constant-volume air thermometer is arranged so that the pressure in the bulb is equal to atmospheric pressure when it is immersed in melting ice. When the bulb is transferred to a vessel filled with steam at 100°C . the pressure is increased by 278 mm. of mercury. What is the temperature of a water bath which would raise the pressure in the thermometer to be 100 mm. above atmospheric? *Camb.**

CHAPTER 17

SPECIFIC HEAT, THERMAL CAPACITY, ETC.

WE are all more or less familiar with milk bills, coal bills and bills including 'labour.' Milk is charged at so much per pint, coal at so much per cwt. (or ton), and labour at so much per hour. All this is easy enough to understand. In a gas bill, however, we really pay for *heat*, measured in a unit known as a *therm*. It is not quite so easy to think of a therm as it is to think of a pint, or a ton, or an hour. However, one therm is equal to 100,000 *British Thermal Units*, and a *British Thermal Unit (B.Th.U.)* is the amount of heat required to raise the temperature of 1 lb. of water 1 degree Fahrenheit.

For scientific purposes the unit most often employed is the *calorie*, which is the amount of heat required to raise the temperature of 1 gm. of water 1 degree Centigrade.

One further point we must mention, because we shall need it almost at once. Just as to raise the temperature of 1 gm. of water 1° Centigrade it is necessary that 1 cal. should be taken in, so when 1 gm. of water cools down 1° Centigrade 1 cal. is given out. Indeed, we could have defined a calorie as the amount of heat given out when the temperature of 1 gm. of water falls 1° Centigrade.

It will be noticed that in these definitions the substance of water is always specified, and perhaps we have asked ourselves whether this really matters. Would lead or sand or mercury for instance, in the process of cooling down, give out the same amount of heat as an equal weight of water?

It is quite easy to put this question to the test of experiment.

Fit a large dry test-tube (6 in. \times 1 in.) with a two-holed cork. Through one hole push a thermometer to within an inch or so of the bottom of the tube, and through the other a bent piece of glass tubing (Fig. 17/1), the longer arm being 6 or 8 in. Weigh out 30 gm. of dry sand and put it in the test-tube.

Now put the test-tube in a beaker of water and boil the latter. The tube serves to prevent moisture from getting into the test-tube. Meanwhile, into a second beaker put 50 c.c. of water from a measuring jar.

When the temperature of the sand has reached, say, 95° remove the cork with its fittings, and then quickly take the temperature of the water.

from the water and empty the contents into the beaker of water. Stir, and notice the highest temperature reached. Repeat the experiment, but this time use 30 gm. of water instead of 30 gm. of sand. In this case the 30 gm. of water will be heated to 95° C. directly over a bunsen, instead of as in Fig. 17/1. Is the rise of temperature produced by adding 30 gm. of water at 95° C., the same as that produced by adding 30 gm. of sand at 95° C.?

Repeat with 30 gm. of lead shot, and then with 30 gm. of mercury. In these cases it will be better to use the method of heating indicated in the next figure.

In one such experiment it was found that on adding the sand the temperature of the cold water was raised from 15° to 22° , while on adding the (hot) water it was raised from 15° to 43° . Clearly the water has had a much greater heating effect than the sand, in spite of the fact that it has cooled down only 52° (from 95° to 43°), while the sand has cooled down 73° (from 95° to 22°). Water, then, in cooling down through a certain range of temperature gives out much more heat than does an equal weight of sand. Correspondingly, if we wish to raise the temperatures of equal weights of water and sand by equal amounts, we must supply far more heat to the water than to the sand. Our experiments with mercury and lead shot should show that the amounts of heat required for these substances are much less than those required either for water and sand.

It will now be clear that the amount of heat required to raise the temperature of a body depends not only on the increase of temperature required and on the mass of the body, but also on the *material* of which the body is made. If the increase of temperature required is 1° C. the amount of heat is known as the 'thermal capacity' of the body, i.e. the **thermal capacity of a body** is the number of calories required to raise the temperature of that body 1 degree centigrade.

It follows from this, and from our definition of a calorie (p. 164), that the thermal capacity of, say, 30 gm. of water would be 30 cals., of 100 gm. of water would be 100 cals., and so on.

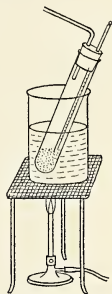


FIG. 17/1

Our experiment showed that sand has a much lower thermal capacity than water. The ratio of the two is known as the 'specific heat of sand,' and similarly for other substances, the **specific heat of a substance** is the ratio of the thermal capacity of that substance, to the thermal capacity of an equal mass of water.

Suppose the thermal capacity of 100 gm. of sand is 19 cal. Now the thermal capacity of an equal mass of water (100 gm) is 100 cal.

$$\begin{aligned}\therefore \text{specific heat of sand} &= \frac{\text{thermal cap. of sand}}{\text{thermal cap. of equal mass of water}} \\ &= \frac{19 \text{ cal.}}{100 \text{ cal.}} \\ &= 0.19.\end{aligned}$$

The thermal capacity of 1 gm. of sand is 0.19 calories. Hence the specific heat of sand or other substance is *numerically* equal to the thermal capacity of 1 gm. of sand, etc. We emphasise 'numerically,' because specific heat is simply a number, a ratio; while thermal capacity is expressed in calories.

Suppose we are dealing with m gm. of a substance of specific heat s , and we raise its temperature by t degrees Centigrade. Then

s cal. will raise temperature of 1 gm. of substance by 1°C

$$\therefore ms \quad \text{,,} \quad \text{,,} \quad \text{,,} \quad m \quad \text{,,} \quad \text{,,} \quad 1^\circ \text{C}$$

$$\therefore mst \quad \text{,,} \quad \text{,,} \quad \text{,,} \quad m \quad \text{,,} \quad \text{,,} \quad t^\circ \text{C}$$

If we call this number of calories h , we have $h = mst$, a very useful relationship.

Example 1. How much heat would be required to raise the temperature of 90 gm. of mercury from 20°C . to 220°C . if the specific heat of mercury is $1/30$?

Here $m = 90$ gm., $s = 1/30$, and $t = 220 - 20 = 200^\circ \text{C}$.

$$\begin{aligned}\therefore h &= mst \\ &= 90 \times 1/30 \times 200 \\ &= 600 \text{ cal.}\end{aligned}$$

When dealing with measurement of heat, we often have to consider the term 'water equivalent.' The **water equivalent of a body** is the number of grams of water which has a thermal capacity equal to that of the body.

Thus if the thermal capacity of a copper calorimeter is 3.2 cal., its water equivalent must be 8.2 gm., because this is the number of grams of water which has a thermal capacity of 3.2 cal. Notice that thermal capacity and water equivalent are *numerically* the same, but the first is expressed in calories, while the second is expressed in grams.

Example 2. The specific heat of aluminium is 0.217. Find (a) the thermal capacity, (b) the water equivalent, of an aluminium calorimeter weighing 100 gm.

We could use the 'mst' formula, but the question is easily worked out from first principles. Thus

To raise temp. of 1 gm. by 1°C . we need 0.217 cal.

„ „ „ 100 gm. „ „ „ $0.217 \times 100 = 21.7$ cal.

∴ thermal capacity of vessel = **21.7 calories**.

The water equivalent is numerically equal to this and is therefore **21.7 gm.**

Water equivalent of a calorimeter. It is often necessary to find the water equivalent of a calorimeter. This may be done as follows:—

If possible, place the calorimeter A inside a larger vessel B, filling the space between (shaded in Fig. 17/2) with felt. This protects A from heat losses.

Pour enough water from a measuring jar to fill A about one-half full, taking the reading before and after.

The difference of course gives the number of c.c. (no. of gm.) in the calorimeter. Take the temperature, estimating to $\frac{1}{10}^{\circ}\text{C}$.

Heat some water in a beaker to about 40°C .

Take the temperature, and then as smartly as possible pour enough of the warm water into the

calorimeter to fill it about two-thirds full. Stir

the thermometer and take the final temperature.

By means of the measuring cylinder¹ find the total amount of water in the calorimeter. The results and calculations appear somewhat as follows:—

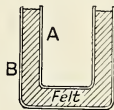


FIG. 17/2

Wt. of cold water = 42 gm.

(measuring jar readings were 100 and 58)

Temp. of cold water . . . 15.2°C .

„ „ warm water . . . 41.7°C .

„ „ mixture . . . 28.8°C .

Or the balance (in which case the weight of the calorimeter must be found). It is quite sufficient to weigh to the nearest decigram.

Total volume after adding warm water = 93 c.c.

\therefore volume of water added = $93 - 42 = 51$ c.c. (= 51 gm.)

Let W = required water equivalent.

51 gm. of warm water have cooled down from 41.7° to 28.8° .

\therefore heat given out = $51(41.7 - 28.8) = 51 \times 12.9$ cal. (a)

42 gm. of water have been heated from 15.2° to 28.8° .

\therefore heat taken in = $42(28.8 - 15.2) = 42 \times 13.6$ cal. (b)

As regards thermal capacity, the calorimeter is equivalent to W gm. of water, and it has been heated from 15.2° to 28.8° .

\therefore heat taken in by calorimeter = $W(28.8 - 15.2)$
 $= W \times 13.6$ cal. (c)

\therefore total heat taken in = (b) + (c) = $(42 \times 13.6) + 13.6W$.

$\therefore (42 \times 13.6) + 13.6W = 51 \times 12.9$

$\therefore 571.2 + 13.6W = 657.9$

$\therefore 13.6W = 86.7$

$\therefore W = 6.4$ grams.

The thermal capacity of the calorimeter would of course be **6.4 calories**.

Specific heat. We are now in a position to find the specific heat of a substance. We will first consider a rough method, then a more exact one.

Rough Method. Arrange a calorimeter with outer case and padding, as already described, and pour into it a measured quantity of water—enough to half-fill it (but see footnote, p. 16).

Tie a piece of cotton round the head of a 50-gm. weight and suspend the latter for two or three minutes in boiling water.

Take the temperature of the water in the calorimeter. Bring the latter near the beaker, and as quickly as possible transfer the weight to the water in the calorimeter. Stir to obtain the final temperature—the highest reached. The result can then be worked out.

Weight of water	= 53 gm.	. . .	(= 53 c.c.)
Initial temperature of water	15.2° C.
" " " brass	100° C. (assumed)
Final " " mixture	21.2° C.
Weight of brass	50 gm.

Let s = required specific heat.

Then heat given out by 50 gm. of brass, of specific heat s , cooling from 100° to $21.2^\circ = 50 \times s \times (100 - 21.2)$ cal. . . . (a)

Heat taken in. As regards thermal capacity, the calorimeter itself is equivalent to 6.4 grams of water (above; we are supposing it is the same calorimeter). We may consider then that the total amount of water in question is $53 + 6.4 = 59.4$ gm.

The temperature of this rises from 15.2° to 21.2° .

Heat taken in = $59.4(21.2 - 15.2)$ cal. (b)

Equating (a) and (b) we have

$$50 \times s \times 78.8 = 59.4 \times 6.0$$

$$\therefore s = \frac{59.4 \times 6.0}{50 \times 78.8}$$

$$= 0.090.$$

The chief sources of error in the experiment are (i) loss of heat in transferring the brass weight, causing the temperature of the water in the calorimeter to rise *less* than it should; (ii) a little hot water is carried with the brass weight, causing the temperature of the water in the calorimeter to rise *more* than it should. These errors often partially cancel each other, and we get better results than we deserve.

By using the apparatus shown in Fig. 17/3 we can completely eliminate error (ii), and greatly reduce error (i).

The calorimeter is a metal cylinder about 8 in. \times 2½ in. It is fitted with a good cork bored to receive a wide test-tube B and the bent tube C.

The metal whose specific heat is required is placed in B preferably in the form of borings, clippings, and is heated by boiling the water in A. Its temperature is measured by the thermometer T.

When T is recording a steady temperature (98° or so) it is removed, cooled under the tap, wiped and used to take the temperature of the water in the calorimeter. (Cf. 'initial temperature of water 15.2° C.' in previous experiment.) To add the metal to the water, A is now grasped with a duster *taking care that C is pointing upwards* it is quickly tipped

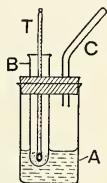


FIG. 17/3

the working out of the result is exactly as before.

The method of finding the specific heat of a substance by heating it, mixing it with water and observing the final temperature, is known as the *method of mixtures*.

Specific heat of a liquid. Sometimes this can be found by the method just described. In many cases, however, heat is made of sheet zinc these cylinders last for a long time. A metal calorimeter will make enough for a class at a very modest cost.

is produced by the mere mixing of the liquid with water *the same temperature*. Alcohol mixed with water is a case in point. In such cases we evidently cannot find the specific heat of a liquid by the method described above. Again, the method would not give good results with liquids such as paraffin, which would not give up their heat readily to the water because they would tend to float on the top of it.

The best plan in such cases is to put the liquid in the calorimeter and to add to it a metal of known specific heat. In the operations are as already described, except that (unless the density is known) we should have to find the quantity of liquid by weighing, not by measuring. Consider the following example, in which we are finding the specific heat of alcohol. The metal added to it is copper, of known specific heat (0.095).

Wt. of calorimeter	64.2 gm.
Calorimeter + alcohol	100.8 "
Wt. of copper added	32.3 "
Initial temp. of alcohol	15.5° C.
" " " copper	98.1° C.
Final " " mixture	24.1° C.

W.E. of calorimeter (previously found) 6.4 gm.

Let s = required specific heat of alcohol.

Wt. of alcohol = $100.8 - 64.2 = 36.6$ gm.

Heat given out by 32.3 gm. of copper, specific heat 0.095 cooling from 98.6° to 24.1° = $0.095 \times 32.3 \times (98.6 - 24.1)$
 $= 0.095 \times 32.3 \times 74.5$ calories.

Heat taken in (i) by 36.6 gm. of alcohol, sp. ht. s , when heated from 15.5° to 24.1° = $36.6 \times s \times 8.6$ cal.

(ii) by calorimeter of W.E. 6.4 gm. when heated from 15.5° to 24.1° = 6.4×8.6 cal.

Total heat taken in = $(36.6 \times s \times 8.6) + (6.4 \times 8.6)$

Now *heat given out* = *heat taken in*.

$\therefore 0.095 \times 32.3 \times 74.5 = (36.6 \times s \times 8.6) + (6.4 \times 8.6)$

$\therefore 228.6 = 314.8s + 55$

$\therefore s = \frac{228.6 - 55.0}{314.8} = 0.55.$

Calorific Value of Fuel. By this time we have become quite used to the idea that heat can be measured. Let us now consider one or two heat measurements which, though they

cannot conveniently be made in an elementary laboratory, so important that we ought to know something about them.

Most people who have used both wood and coal as fuel would agree that wood gives out less heat than an equal weight of coal, and this leads us to the question of what is called the *calorific value* of fuel, by which we mean the heat given out when one gram of the fuel in question is completely burnt. Suppose a certain kind of coal is being examined in this way. About a gram of a sample is finely powdered and pressed into the form of a cylindrical pellet. The latter is then accurately weighed and placed in a special 'bomb calorimeter' designed to stand high pressure. Oxygen at ten atmospheres pressure is passed into the bomb, which contains a thin wire capable of being electrically heated.

The bomb is immersed in a vessel containing a known weight of water. On passing the current for a moment the pellet takes fire and is allowed to burn itself out, the heat evolved passing into the surrounding water. Suppose this is 1000 gm., and that the temperature rises from 12° to 18° C. Suppose also that the weight of coal is 1 gm. and that the water equivalent of the calorimeter is 100 gm. Then (omitting certain minor corrections) 1 gm. of coal when completely burnt gives $1000(18 - 12)$ cal. to the water, and $100(18 - 12)$ cal. to the calorimeter, a total of $6000 + 600 = 6600$ cal. This would be the calorific value of the fuel in question.

The calorific value of all sorts of *food* has been found in the same way. 'But surely we don't use food as fuel?' you say. Well, in a sense we do. In our bodies it undergoes a kind of slow burning, and except for two important qualifications, it produces the same amount of heat as if it were burnt in a bomb calorimeter. The two qualifications are these.

(1) The body is not able to burn up the *whole* of the food, though in many cases (starch, sugar, fats) the percentage burnt may be as much as 95 or more.

(2) Very often the heat produced does not appear as such, but as mechanical energy. When we walk, run, jump, etc., we are using energy which is the *equivalent* of heat, and this energy is derived from food. You will understand this better by studying the Mechanical Equivalent of Heat. in

Chapter 21. A man doing heavy physical work needs as much food per day as will produce from 3400 to 4000 'great calories' (1 great cal. = 1000 cal.). If he is doing only light work about three-fourths of this amount is sufficient. The all-round average for the entire population should be about 3000.¹

Determination of high temperatures. When we are finding the specific heat of a solid such as copper by mixing hot copper with cold water, we know the temperature to which the metal was raised, and can calculate the specific heat. Conversely, if we know the specific heat, we can calculate the temperature.

This principle was used to some extent at one time, for finding the temperature of furnaces, etc., the metal used being platinum. The great weakness of the method is that the temperature of the mass of platinum falls very rapidly when once it is removed from the furnace, and hence the value of t obtained is much too low. High temperatures are not recorded quickly and accurately by means of electric resistance pyrometers, the action of which depends on the fact that the resistance of a metal wire varies with the temperature.

Questions

A

1. How many British thermal units would be required to raise the temperature of a gallon of water (= 10 lb.) from 62° to the boiling point?

2. How many calories would be needed to raise the temperature of (i) 20 gm. of water from 10° C. to 60° C., (ii) 1 kgm. of water from 20° C. to 30° C.?

3. What is the thermal capacity of (i) a penny, assuming it weighs 9 gm. and to be made of material whose specific heat is 0.095, (ii) a silver tea-pot weighing 400 gm. (sp. heat of silver = 0.057).

4. What would be the water equivalent of the teapot of q. 3?

¹ These appear to be the latest 'official' figures and were put forward in 1934 by an expert committee of the British Medical Association in conjunction with representatives of the Advisory Committee on Nutrition of the Ministry of Health. The all-round average in November 1947 was about 2700.

B

. Define *specific heat*.

10 gm. of water at 36.3°C . are poured into a calorimeter containing 50 gm. of water at 15.0°C . The final temperature obtained is 24.0°C . Calculate the water equivalent of the calorimeter.

The calorimeter is dried and 100 gm. of a different liquid are put into it. The temperature is then 15.3°C . A piece of metal weighing 50 gm. which has been heated to 100°C . and whose specific heat is 0.2, is placed in the liquid and the temperature rises to 26.8°C . What is the specific heat of the liquid? *Camb.*

. A mass of turpentine at 60°C . and an equal mass of water at 20°C . are mixed and well stirred. Calculate the final temperature, being given that the specific heat of turpentine is 0.2 cal. per gm. per deg. C. Neglect any loss of heat during mixing. *Lond.**

. Define the terms calorie, specific heat, water-equivalent. Describe how you would determine the water-equivalent of a calorimeter by experiment.

A calorimeter whose water-equivalent is 8 gm., contains 80 gm. of oil at 12°C . When 100 gm. of a metal (of specific heat 0.09) at a temperature of 98°C . is placed in the oil, the resulting temperature is 23°C . Calculate the specific heat of the oil. *Camb.*

. A calorimeter weighing 200 gm. contains 50 gm. of water at 20°C . 100 gm. of water at 80°C . is poured into the calorimeter and the final temperature on mixing is 50°C . What is the specific heat of the calorimeter material? *Dur.**

In a test on a high explosive, 0.100 gm. of the explosive is burnt slowly in a thick iron vessel, of mass 2700 gm. and specific heat 0.111, which was immersed in 400 gm. of water. The rise in temperature of the iron and water, allowing for heat losses, was 0.101°C ., find the number of calories which would be produced by 1 gm. of the explosive. *Camb.**

A metal forging of mass 3 lb. is heated to a temperature of 600°F ., and is then quenched by plunging it into oil at 60°F . contained in a vessel whose water equivalent is 2 lb. What is the minimum mass of oil that should be used so that its temperature shall not rise above 80°F .?

Specific heat of metal = 0.1. Specific heat of oil = 0.5.) *Dur.**

Describe *one* method of measuring each of the following, and in each case explain the principle of the instrument or method used: (a) temperature, (b) quantity of heat.

Equal volumes of copper and gold of specific gravities 8.9 and 19.3 respectively have their temperatures raised through 10°C . Calculate the quantities of heat absorbed by the copper and gold are in the ratio 4/3. Compare the specific heats of copper and gold. *O.C.*

CHAPTER 18

LATENT HEAT

EVERYBODY seems to know that when a kettle of water has reached the boiling point it is useless to leave it any longer on the fire, or gas-ring, in the hope that it will become hotter still. It would simply remain at the same temperature (actually 100°C.) until it was all converted into steam. Many people, however, who have quite clear ideas of what happens when heat passes into boiling water, are much less certain of the effect of passing heat into *ice*. Actually, if it is below 0°C. to begin with, its temperature will rise until its melting point (0°C.) is reached, and will then remain constant until the ice has been completely converted into water.

The same holds good with regard to any other substance when at its melting point or boiling point, as the case may be.

In the present chapter we shall particularly study the 'constant temperature' states, especially the constant temperature maintained by ice during its conversion into water and that maintained by boiling water while it is being converted into steam.

To begin with, however, let us try to get some idea of what happens during the heating of a solid.

In some of your P.T. drill you are standing in a particular position among your fellows, and although you make many bodily movements you do not leave the position. The molecules of a solid are something like that. They move and vibrate—but only to one side or other of a fixed central position. If the solid is heated, the average energy of the molecules is increased, for heat is a form of energy (a point we shall consider in greater detail in Chapter 21). Now the kinetic energy of a molecule is represented by the product $\frac{1}{2}mv^2$, where m is the mass and v the velocity (p. 45). The mass m is constant, so an increase in kinetic energy means an increase in v .

In short, if we think of the molecules of a solid as being something like a group of boys 'spaced out' at P.T., then we must think of them as moving more and more violently when

solid is heated—but still keeping close to centres fixed each one. The boys at P.T. keep their relative positions by taking care. The molecules, of course, do not—y keep them because of the attractive force of neighbouring molecules.

But as the temperature rises and the molecules vibrate with ever-increasing velocity, a point is reached when they vibrate so strongly as to break away from the attractions of their neighbours. There is an end to the 'spaced-out' arrangement. The molecules glide in and out among their neighbours something like the individual lead shots in a barrel of lead shot that is being stirred.¹ We say that the solid has *melted*.

Note that while the solid is being melted, the energy we are supplying in the form of heat is being used not to increase the rate of vibration of the molecules, but to enable them to break away from the attractive forces holding them together. In other words, the heat supplied is not causing an increase in the *kinetic* energy of the molecules, but in their *potential* energy. It is because the kinetic energy is for the time being at a standstill that the temperature remains constant.

Suppose we continue to apply heat after the process of melting is complete. Once more we begin to increase the kinetic energy of the molecules. They glide in and out with greater speed, and no doubt they vibrate with greater intensity at the same time.

Molecules near the surface will, in general, be kept within the liquid by the combined pulls of the molecules in their neighbourhood. The molecules, however, are travelling at different speeds, and if a fast-moving one at the surface happens to be travelling in an upward direction, it is likely to break away from the liquid altogether. This breaking away of molecules from the surface is *evaporation*. As it is the fastest-moving molecules which escape, the average speed of the remainder is reduced, and this is represented by the cooling effect which always accompanies evaporation (*Junior Physics*, pp. 166–8).

The analogy is not very perfect, for while each shot would have a *translatory* motion, the molecule has both a translatory motion and a *vibratory* one.

During evaporation the only molecules that can break away are those that happen to be travelling at comparatively high speed when they are near the surface. But if the liquid is heated, bubbles after a while begin to be formed inside the liquid, and evaporation now takes place not only from the free surface, but into the bubbles, from their inner surface. This new state of affairs, in which the interior of the liquid is concerned as well as the free surface, is known as *boiling*, and evaporation is now much more rapid because it is taking place from a greatly extended surface.

Suppose a flask of water is boiling over a bunsen burner and we then introduce an extra bunsen. What happens?

The existing bubbles are enlarged, and more bubbles are formed, increasing the (internal) surface at which evaporation takes place. Thus we have an increased rate of evaporation. But, as we have already seen, evaporation is accompanied by loss of heat, and the increase in this rate of loss just balances the gain of heat from the extra bunsen. That is why the temperature does not rise.

Latent heat of fusion. When ice is melting, the heat passing into it is said to be *latent* (= hidden), because it cannot be detected by the thermometer. The *latent heat of fusion of ice* is the number of calories required to convert 1 gram of ice at 0°C . into water at the same temperature. You can, no doubt, modify this definition for yourself, so as to make it apply to solid substances in general.

Instead of 'latent heat of fusion of ice,' the shorter term 'latent heat of ice' is often used. The method of finding the latent heat is seen in the following experiment:—

Take a double-walled calorimeter (Fig. 17/2, p. 167), weigh the inner vessel (to one decimal place only; and so for all weighings in this experiment). Half fill the calorimeter with water and weigh again.

Using a small bunsen flame, warm the calorimeter until the temperature of the water is about 6 degrees above that of the room, and then place it inside its padded enclosure.

Now break up some ice into pieces about the size of peas, dry them by rolling them over on a sheet of blotting-paper. A suitable quantity would have, roughly, one-eighth the weight of the water in the calorimeter.

Take the temperature of this water (reading to 0.1°C .) and quickly add the ice, picking up the latter by means of crucibles.

s (preferably with blotting-paper tied round the tips). Stir the time with the thermometer, and record the lowest temperature reached. Weigh again so as to find out how much ice has been added. The results are then worked out as follows:—

Suppose (a) Wt. of calorimeter	62.3 gm.
(b) Calorimeter + water	133.6 „
(c) Calorimeter + water after addition of ice	142.4 „
Initial temperature	21.3° C.
Final temperature	11.2° C.
W.E. of calorimeter (p. 168)	6.4 gm.

Weight of ice used = (c) - (b) = 8.8 gm.

Let l = required latent heat.

Heat required to melt 8.8 gm. of ice = $8.8l$ cal.

Note that $8.8l$ cal. suffices only to *melt* the ice, i.e. to turn it into water at 0° C.) In our experiment this melted ice is raised to a temperature of 11.2° C., and this requires a further quantity of heat equal to (8.8×11.2) cal.

Total amount of heat required to convert 8.8 gm. of ice at 0° into water at 11.2° C. = $8.8l + (8.8 \times 11.2)$ cal. . (A)

This heat has been supplied by the cooling of the water in the calorimeter ($133.6 - 62.3 = 71.3$ gm.) from 21.3° to 11.2°, and by the cooling of the calorimeter itself (= 6.4 gm. of water).

Total weight of water may be taken as $71.3 + 6.4 = 77.7$ gm.

Fall of temperature = $21.3 - 11.2 = 10.1^\circ$.

Heat given out = 77.7×10.1 cal. (B)

Equating (A) and (B) we have

$$8.8l + (8.8 \times 11.2) = 77.7 \times 10.1$$

$$\therefore 8.8l + 98.6 = 784.8$$

$$\therefore 8.8l = 686.2$$

$$\therefore l = 78.0 \text{ cal. per gm.}$$

One or two points connected with this experiment call for a little discussion.

Weighing only to one place of decimals. We cannot read temperatures more accurately than to one place of decimals, and it is useless to take weighings to two places. 'A chain is no stronger than its weakest link.'

Preliminary heating of water. Suppose this is heated to 21°, and cools down to 11°, the temperature of the room being 16°.

We may divide the cooling process into two parts: (a) from 21° to 16° and (b) from 16° to 11°.

During (a) the water is *warmer* than the air in the room and so loses some heat to the air.

During (b) the water is *colder* than the air, and so gains heat from the air.

If the initial temperature is just as much above that of the room as the final temperature is below it (as in the case under discussion) the two errors will cancel out. To find exactly how much to warm the water, a preliminary experiment is evidently necessary.

(iii) *Drying the ice and using padded tongs.* In our calculations we assume that no ice is melted except by heat obtained from water in the calorimeter. Evidently this is true only if the ice is dry when it enters the calorimeter.

(iv) *Padding the calorimeter.* This, of course, is common to nearly all experiments in which a calorimeter is used, and is intended to minimise loss of heat to the air and the supporting table (or equally, to minimise gain of heat).

Latent heat of vaporisation. We have seen that it takes a great deal of heat—actually 80 cal.—simply to melt 1 gm. of ice without producing any rise of temperature. Similarly, when water is boiling it takes a great deal of heat simply to turn 1 gm. of it into steam without rise of temperature. This heat is known as the *latent heat of vaporisation of water*, or more shortly, as the *latent heat of steam*, and may be formally defined as *the number of calories required to convert one gram of boiling water¹ into steam at the same temperature.*

A method² now much used for finding the latent heat of steam involves the use of a heavy calorimeter.

The apparatus is represented in Fig. 18/1. Water is boiled in the tin vessel A (or a flask might be used) and passes into the heavy metal calorimeter B, which is surrounded by a felt cover. B has previously been weighed on a 'household balance.' Steam can escape through the bent tube C.

To begin with, the calorimeter B is at the temperature of the room, which is noted. When steam begins to enter, it condenses, giving up its latent heat to B, the temperature of which is raised. Once B has attained a temperature of 100° C. it is unable to

¹ The boiling point varies according to the pressure, and so strictly speaking, there is a special latent heat for each boiling point. We shall confine ourselves, however, to the case of the boiling point at normal pressure, which for water would be 100° C.

² Devised by Mr. A. R. Marshall, H.M.I.

condense any more steam, and the latter then begins to escape through C. When it is escaping freely the bunsen is turned out, and the inverted calorimeter with its cork and two tubes is moved. The felt is taken off, and the calorimeter is held under a tap, to cool the contained water. The latter is then measured either in a cylinder or (better) by pouring it through a funnel into a burette. The specific heat of the metal of which the calorimeter is made must be known, and the result is then worked out as follows:—

Weight of calorimeter	1142 gm.
Specific heat (brass)	0.092
Temperature of room	15.4° C.
Wt. of steam condensed (16.8 c.c.)	16.8 gm.

Let L = required latent heat.

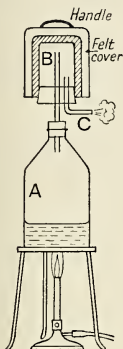


FIG. 18/1

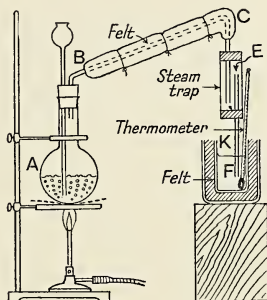


FIG. 18/2

When heat given out by condensation of 16.8 gm. of steam
 $= 16.8L$ cals.

Heat taken in by calorimeter as its temperature rises from
 15.4° to 100°

$$= 1142 \times 0.092(100 - 15.4) = 1142 \times 0.092 \times 84.6 \text{ cal.}$$

Equating, we have $16.8L = 1142 \times 0.092 \times 84.6$

$$\text{giving } L = 529$$

The method gives very fair results. The chief source of error is in connection with 'weight of steam condensed.' The above typical result shows that an error of 1 gm. in this would throw the result out by about 30 units. Some is apt

to be lost by evaporation, and by clinging to the sides of the jar or burette.

The container A should not be more than, say, a third full; otherwise, during boiling, water is apt to be thrown into the calorimeter.

Another method. An older method, in which steam is passed into cold water, is still widely used.

The apparatus of Fig. 18/2 is fitted up. Water is boiled in the flask A, and the steam passes through the tube BC, which is packed round with felt. BC is given a backward slope so that condensed steam may run back into the flask. Some condensed steam is also caught in the steam-trap (really a water-trap). From the latter an exit tube EF passes into the calorimeter, which is contained as usual in an outer vessel lined with felt or cotton-wool.

The calorimeter is first weighed empty, then about two-thirds full of water (one decimal place). Before placing the calorimeter in position, the water in A is allowed to boil freely for two or three minutes so that all tubes, etc., may be thoroughly heated. This greatly reduces the amount of steam condensed *en route* to the calorimeter.

The temperature of the cold water in the calorimeter is now taken, and steam is passed in until the temperature has risen 30° or so. The calorimeter is then withdrawn and the final temperature taken. It is then weighed *quickly*, because the warm water contained in it evaporates rather quickly. Suppose

Weight of empty calorimeter.	62.3 gm.
Calorimeter + water	153.4 „
Calorimeter + water + condensed steam	158.5 „
Initial temperature	15.7° C.
Final temperature	45.9° C.
W.E. of calorimeter	6.4 gm.

Let L = required latent heat. As L calories are required to convert 1 gm. of water at 100° into steam, we shall assume that when 1 gm. of steam at 100° condenses to water (still at 100°) it gives out L cal.

In our experiment, wt. of steam condensed

$$= 158.5 - 153.4 = 5.1 \text{ gm.}$$

\therefore heat given out during condensation = $5.1L$ calories.

The condensed steam then cools from 100° to 45.9° , giving a further $5.1(100 - 45.9) = 5.1 \times 54.1$ cal.

\therefore total of heat given out = $5.1L + (5.1 \times 54.1)$ cal. (A)

Wt. of water in calorimeter = $153.4 - 62.3 = 91.1$ gm., and the calorimeter itself is equivalent to 6.4 gm.

\therefore wt. of water may be taken as 97.5 gm.

Temperature of this water is raised from 15.7° to $45.9^{\circ} = 30.2^{\circ}$

\therefore heat taken in by water = 97.5×30.2 cal. (B)

Equating (A) and (B) we have

$$5.1L + (5.1 \times 54.1) = 97.5 \times 30.2$$

Solving, we obtain $L = 524$ calories per gram.

The chief sources of error are: (i) evaporation after passing steam (already mentioned); (ii) the increase in weight should be due entirely to *steam* entering the calorimeter. Actually some water enters (due to premature condensation). This error is reduced by making the tube EF as short as possible.

We will conclude this chapter with two numerical examples.

Example 1. An aluminium kettle weighs 100 gm. and 1 litre of water is put into it. Its temperature is 15°C . It is then placed on a gas ring and in 10 minutes the water boils. If heat continues to be supplied at the same rate, how long will it be before all the water is away? (Specific heat of aluminium = 0.22 and latent heat of steam = 540.)

- (i) Heat required to raise temperature of 100 gm. of aluminium from 15°C . to 100°C . = $0.22 \times 100 \times (100 - 15)$
 $= 0.22 \times 100 \times 85 \text{ cal.} = 1870 \text{ cal.}$

Heat required to raise temperature of 1 litre (= 1000 gm.) of water from 15°C . to 100°C .

$$= 1000 \times 85$$

$$= 85,000 \text{ cal.}$$

Total amount = $85,000 + 1870 = 86,870$ calories.

- (ii) Heat required to evaporate 1 gm. of water = 540 cal.

\therefore heat required to evaporate 1000 gm. = 540,000 cal.

From (i) we see that the gas-ring supplies 86,870 cal. in 10 minutes = 8687 cal. per minute.

In (ii), 540,000 cal. are required.

$$\therefore \text{ time required} = \frac{540,000}{8687} = 62.2 \text{ minutes.}$$

Example 2. 12 gm. of steam are blown into a large block of ice. How much ice would be melted? (Latent heat of steam = 540; specific heat of ice = 80.)

In the process of condensing, 12 gm. of steam will give out $12 \times 540 = 6480$ cal.

In the further process of cooling down from 100° to 0° , heat given out by water = $12 \times 100 = 1200$ cal.

Total amount of heat given out = $6480 + 1200 = 7680$ cal.

Now for every 80 calories given out, 1 gm. of ice is melted.

$$\therefore \text{ no. of gm. of ice melted} = \frac{7680}{80} = 96.$$

Questions

A

1. How many calories would be needed (i) to raise the temperature of 10 gm. of ice from -20°C. to the melting point (ii) to melt the ice; (iii) to raise the temperature of the ice-water to 100°C. ; (iv) to evaporate the water? (Specific heat of ice = 0.5; the latent heat of fusion of ice = 80 cal. per gm. latent heat of vaporisation of water = 540 cal. per gm.)

2. 100 gm. of steam at 100°C. are blown into the middle of a quantity of ice at 0°C. How much ice would be melted? (Latent heats of fusion and vaporisation as given in q. 1.)

3. 10 gm. of ice at 0°C. are dropped into 50 gm. of water at 20°C. Find the final temperature. (Latent heat as in q. 1.)

4. 10 gm. of steam at 100°C. are passed into 100 gm. of water at 20°C. Find the final temperature. (Latent heat as in q. 1.)

B

1. Describe simple experiments, one for each case, to illustrate the meaning of the terms *latent heat*, *specific heat*.

15 gm. of dry ice at 0°C. are dropped in 200 gm. of liquid of specific heat 0.52 cal. per gm. per deg. C., contained in a calorimeter of water equivalent 15 gm. If the original temperature of the liquid was 30°C. , find the temperature when the ice has just melted. (The latent heat of fusion of ice is 80 cal. per gm.) *Lond.*

2. Define *latent heat of fusion* of a solid. Describe how you would measure the latent heat of fusion of ice.

What mass of ice at 0°C. must be mixed with 500 gm. of water at 17°C. to lower its temperature to 10°C. ? The water equivalent of the vessel containing the water is 13 gm. (The latent heat of fusion of ice is 80 cal. per gm.) *Lond.*

3. A thermos flask whose water equivalent is 30 gm. contains 500 gm. of water at 20°C. What is the least amount of ice required to cool the water and flask to 0°C. ? *Dur.*

4. Describe an experiment to show that water expands on solidification. Give *two* examples of other substances which also expand on solidification.

During a storm, a depth of 4 cm. of rain falls on the surface of a frozen pond. The temperature of the rain is 8°C. and that of the ice 0°C. Find the thickness of ice melted, given that the S.G. of ice = 0.92, and the latent heat of fusion of ice = 80 calories per gram. *O.C.*

5. 50 gm. of mercury are frozen in a block of solid carbon dioxide to -80.0°C. and transferred to a calorimeter (of water equivalent 4 gm.) containing 36 gm. of water. The temperature falls from 17.5°C. to 10.0°C. Taking the specific heat of mercury in both the solid and liquid states to be 0.033, calculate its latent heat of fusion. *Camb.**

. A calorimeter of water equivalent 10 gm. contains 100 gm. water and 10 gm. of ice. How much steam must be passed to the calorimeter to raise the temperature to 50°C .? (Latent heat of ice = 80 calories per gm.; latent heat of steam = 540 calories per gm.) *Dur.**

. Suggest an experiment to show that water can be frozen by its own evaporation. What weight of water at 0°C . must evaporate in order to freeze 1 gm. of the remaining water? (Latent heat of fusion of ice at 0°C . = 80 cal. per gm.; latent heat of evaporation of water at 0°C . = 606 cal. per gm.) *Camb.**

. What do you understand by the terms: the *thermal capacity* of a body, the *latent heat of vaporisation* of a substance, *melting point*?

. A copper calorimeter of water equivalent 10 gm. contains 10 gm. of water and 100 gm. of ice, all at 0°C . Heat is supplied to it electrically at the rate of 500 cal. per sec. How many seconds from the start will it take for (a) all the ice to melt, (b) the calorimeter and contents to be at 100°C ., (c) the calorimeter to have boiled dry? (Latent heats of water and steam = 80 and 540 cal. per gm. respectively.) *Camb.*

. Explain why a cooling effect accompanies the evaporation of liquids and give *two* examples. Describe an experiment in which water is frozen by evaporation.

. Calculate how much steam at 100°C . will melt 10 gm. of ice at 0°C . if the water produced remains at 0°C . Take the latent heats of fusion and evaporation of water as 80 cal./gm. and 540 cal./gm. respectively. *O.C.*

. Give an account of the mode of action of a steam engine, and illustrate your answer with a large clear diagram.

. Steam is generated by means of a gas heater, find the cost of changing 1 gallon of water at 40°F . into steam at 212°F ., given that 55 per cent. of the heat supplied by the combustion of gas is utilised in heating the water and changing it to steam, and that the cost of the gas is fourteen pence per therm. (1 gallon of water weighs 10 lb. 1 therm = 10^5 British Thermal Units. Latent heat of vaporisation of water = 972 British Thermal units per lb.) *C.W.B.*

CHAPTER 19

VAPOUR PRESSURE

WE all know that when water evaporates it disappears completely—water vapour is just as invisible as air itself. The air round about us always contains more or less of this invisible vapour. We know, too, that important results follow from its presence—that somehow it is connected with such subjects as mist, fog, dew and other weather phenomena. Let us try to turn our rather hazy knowledge of water vapour and the part it plays into something fuller and more exact.

Saturated and Unsaturated Vapour. We can learn a great deal from a few simple experiments with a barometer tube. A tube AB about a yard long is completely filled with mercury and inverted in a dish of mercury. The column of liquid falls a little, leaving a vacuum in the space AC, while the length DC is 'the height of the barometer.' It is a good plan to mark the position of C with gummed paper.

Some water is now boiled well to free it from air, and after cooling, a very tiny quantity is passed upwards through the mercury into the space AC. This is best done by means of a fountain-pen filler, the end of which has been bent in a gas-flame.

If the quantity of water introduced has been sufficiently small it will vaporise completely on entering the vacuum, and the mercury will be pushed down a millimetre or two to the level G (Fig. 19) in which, however, the mercury must be supposed to rise upwards as far as G. The length CG (say 1.5 mm. of mercury) is a measure of the vapour pressure.

On introducing a further tiny quantity of water the mercury is pushed down a little more, but a stage is very soon reached at which the space is 'saturated' with vapour. More water may be introduced, but it simply rests unvaporised on



FIG. 19

of the mercury (EF in Fig. 19/2), while the mercury level is no longer depressed. We are supposing that we have been working at 'room temperature,' say 15°C . throughout. In that case, CE would measure 12.7 mm.,¹ and we say that the maximum pressure of aqueous vapour at 15°C . is 12.7 mm.—'Maximum,' because at 15°C . water vapour can exert any pressure between zero and 12.7, according to the amount present. When the space is saturated, the depression reaches its maximum value, 12.7 mm.

If we compress a given quantity of *air* into a smaller space, we increase its pressure (Boyle's law). If, however, we compress our *saturated vapour* into a smaller space,² all that happens is that some of it condenses and the pressure remains constant at 12.7 mm. If we increase the volume of the space enough of the surplus water vaporises to keep the space saturated and to keep the pressure at 12.7 mm., until, of course, the space becomes so large that all the surplus water is vaporised; beyond that point the space becomes unsaturated, and the pressure begins to fall.



FIG. 19/2

Contrast all this with the case of an unsaturated vapour in a confined space. As we increase the space, the vapour pressure becomes less. As we decrease it, the pressure increases up to the point at which the space is saturated. Do other liquids, at 15°C ., exert a maximum vapour pressure of 12.7 mm.? The question is easily answered by repeating the whole experiment, but using alcohol instead of water. A maximum pressure of 33.0 mm. is obtained. With ether it is 354 mm. Sulphuric acid, on the other hand, gives a vapour pressure of less than 1/100 mm. Evidently this maximum pressure, at a given temperature, varies a great deal from one liquid to another.

So far all our experiments have been carried out at 15°C ., whatever 'room temperature' happened to be. Now, going back to the barometer tube with a little unvaporised water and the mercury (Fig. 19/2), let us warm the space AE a little. It is not suggested, however, that such an accurate result could be obtained with the simple apparatus described. Easily done by pushing the barometer tube farther down into the reservoir of mercury.

little by waving a bunsen flame over it. Some of the water vaporises, and the mercury is at once pushed farther down. Evidently *the maximum pressure of aqueous vapour increases with temperature.*

Much more exact experiments can be carried out with the apparatus shown in Fig. 19/3. Here the barometer tube is graduated, and is enclosed in a wider glass tube through which water from a geyser can be circulated. A thermometer through the upper cork gives the temperature.

Suppose we wish to find the maximum pressure of aqueous vapour at 40°C . We regulate the flow of water from the geyser until the temperature is as steady as possible at 40° . We take the barometer reading, pass in some water with the curved filler and take the reading again (making sure, of course, that there is a little unvaporised water above the mercury). Suppose the readings are respectively 762 mm. and 707 mm. Evidently the required vapour pressure is 55 mm.

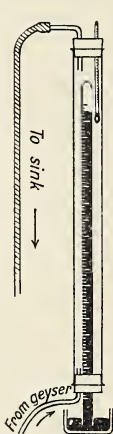


FIG. 19/3



FIG. 19/4

An important special case is the pressure¹ of aqueous vapour at the boiling point. Here we must discard the geyser, and connect the upper tube to a flask generating steam, which passes round the barometer tube and escapes through the lower tube (then, preferably, through a condenser).

When the steam has been passing a little time it is found that the mercury in the barometer tube is pushed down (Fig. 19/4)—i.e. to the level of the mercury in the outer tube. The pressure inside the tube is therefore equal to that of the atmosphere. A similar result is obtained for other liquids.

¹ 'Maximum' is understood unless the contrary is stated.

g. the pressure of ether vapour at 35° (this being the boiling point of ether) is equal to that of the atmosphere.

N.B. In the last experiment it would be easier, and much safer, to let water at 35° circulate through the wide tube, than to make use of vapour from boiling ether.

We may therefore state generally that *the maximum vapour pressure of any liquid, at its boiling point, is equal to the pressure of the atmosphere.*

Relative Humidity. We often talk about the air being 'damp.' We perspire, and the perspiration remains on our bodies instead of drying off. At such times a line of washing may take many hours to dry, although no rain is falling. At other times we talk about the air being 'dry.'

It is important to notice, however, that the sensation of dampness or dryness depends not so much on the amount of moisture actually present, as on what is called the **Relative Humidity**, this being defined as *the ratio of the amount of water vapour present in a given volume of air, to the amount required to saturate that volume of air at the same temperature.*

For instance, 1 cubic yard of air at 20° C. requires 13.0 gm. of water vapour to saturate it, while if the temperature is 10° C., only 7.1 gm. would be required. Suppose the amount of water vapour actually present is 6.5 gm. Then if the temperature is 20° C., relative humidity

$$= \frac{\text{amount present}}{\text{amount required to saturate}} = \frac{6.5}{13} = 0.5, \text{ or } 50\%.$$

If the temperature is 10° C., relative humidity $= \frac{6.5}{7.1} = 0.92$ or

92%. The air would feel very much 'damper' in the second case than in the first, because, although it contains only the same amount of water vapour, it is nearer to saturation point. For many purposes it is important to be able to determine the relative humidity of the atmosphere. It is needed, for instance, in making forecasts of the weather, in deciding whether the atmosphere of a cotton-mill is suitable for spinning and weaving, and in many other cases.

The method of determining relative humidity is indicated in Fig. 19/5 (only roughly, for there are various refinements to be considered here. Ignore the flask F for the present). It consists of a tube containing some substance which will readily

absorb water vapour—calcium chloride, for instance. It is connected with a large bottle or aspirator B. A is weighed and then air is drawn slowly through it by turning on a slow stream of water at C. Suppose that we have run out 5 litres of water, and that the weight of A has increased 0.045 gm. Then the weight of water vapour present per litre = $0.045 \div 5 = 0.009$ gm. Suppose the temperature is 20°C . Reference to tables shows that at this temperature 1 litre of air if saturated with water vapour would hold 0.017 gm.

\therefore relative humidity = $0.009 \div 0.017 = 0.53$, or 53%.

Instead of referring to tables, it is instructive to repeat the experiment with the addition of, say, three flasks (or better

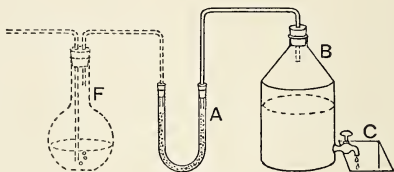


FIG. 19/5

Winchester quart bottles) containing water, of which one, F, is shown in the diagram. The same volume of water is run out as before. The idea is, of course, to find what weight of water vapour is required to *saturate* the given volume of air, but the results obtained are usually too low, making the relative humidity come out too high.

The processes just described are evidently very tedious and much shorter methods are in use depending on a determination of the **dew-point**, which is *that temperature at which the moisture actually present in the air would be sufficient to saturate it*.

Think of a cubic yard of air at 20°C . containing 7.1 gm. of water vapour. That is much less than the amount (100 gm.) which would be required to saturate it. Now let the air be cooled down, the amount of water vapour in it remains constant at 7.1 gm. The vapour-holding capacity of the air becomes less and less, until at a certain temperature

ually 10°C .—the 7.1 gm. of vapour is sufficient to saturate

At this temperature the air begins to deposit its moisture in the form of dew. We shall see presently how to find the dew-point (10°C . in the case just considered).

Now, from tables we can find the amount of water vapour required to saturate 1 cubic yard of air at any given temperature, and we know that

$$\text{Relative humidity} = \frac{\text{weight actually present per cu. yd.}}{\text{weight required to saturate 1 cu. yd.}}$$

The numerator of this fraction is 7.1 gm., for the 'weight actually present' did not change on cooling until the dew-point was reached, and this amount was just sufficient to saturate the air at the dew-point. The denominator is 13.0.

$$\text{Hence Relative humidity} = \frac{7.1}{13} = 0.53 \text{ or } 53\%.$$

Notice that so long as the temperature is constant, the weight of a gas present in a given volume, say 1 cubic yard, is proportional to the pressure. Hence in the fraction given above we may read 'pressure' instead of weight, and we get relative humidity

$$\begin{aligned} &= \frac{\text{actual pressure of aqueous vapour at existing temp.}}{\text{maximum pressure of aqueous vapour at existing temp.}} \\ &= \frac{\text{maximum pressure of aqueous vapour at dew-point}}{\text{maximum pressure at existing temperature}} \end{aligned}$$

Example. Find the relative humidity of air at 15°C . if the dew-point is found to be 8°C . (Maximum pressure of aqueous vapour at 15°C . is 12.67 mm. and at 8°C . is 7.89 mm.)

$$\begin{aligned} \text{R.H.} &= \frac{\text{Maximum pressure of aqueous vapour at } 8^{\circ}\text{C.}}{\text{Maximum pressure at } 15^{\circ}\text{C.}} \\ &= \frac{7.89}{12.67} = 0.623 \text{ or } 62.3\% \end{aligned}$$

How to find the dew-point. About two-thirds fill a bright aluminium beaker with water. Add a bit of ice from time to time, stirring with a thermometer after each addition. Avoid breathing too near the beaker. Observe the temperature at which the metal surface begins to become misty, say 8°C . Now let the vessel gradually warm up, and observe the temperature at which the mist begins to disappear, say 9° . The dew-point may be taken as 8.5° , the mean of the two readings.

An instrument used for finding the dew-point is known as a **hygrometer**. The simple hygrometer just described does not give very accurate results, though it is good for giving us a general idea of the subject. A much better instrument is Regnault's hygrometer (Fig. 19/6). This consists of a sort of test-tube, the lower part of which is made of thin silver, highly polished on the outside. The tube contains ether, which is rapidly evaporated by blowing air through it. The evaporation produces cooling, and the temperature at which a deposit appears is recorded by a thermometer dipping into the ether.

In modern forms of the instrument, the apparatus just described is enclosed in a box. The silver portion is brightly illuminated by means of an electric light, and the observations are made through a window in the side of the box. In this way local variations of humidity due to the observer's breath are prevented.

Conditions for heavy dew. When the temperature of the atmosphere falls to dew-point, dew will begin to be deposited. The amount of this deposit will depend on circumstances.

We are more likely to have a heavy fall of dew after a *hot day*, because the air will have taken up more water in the form of vapour. Next, it is desirable that the earth's surface should be able to cool as completely as possible by radiation, and for this we require a *clear sky*.

The air lying close to the earth's surface must remain in that position while it is cooling down to the dew-point, so we require *absence of wind*.

Again, a solid body which is a good conductor may become very cold, because the heat lost by radiation is constantly replenished by heat conducted to it from the earth below. Such a body is less likely to receive a deposit of dew. For instance, there may be dew on a wooden fence but none on some iron railings close by.

Summing up, we may say that the best conditions for a copious deposit of dew are a calm night with a clear

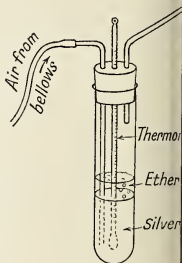


FIG. 19/6

wing a hot day; and the dew will be deposited on non-conducting materials more than on conducting ones.

may happen that the dew-point is below 0° C. In that the water vapour in the air will be condensed in the solid, *i.e.*, as *hoar frost*. If there is no such deposit, though temperature is below freezing point, we speak of a *black*

weating' of Walls. Recently built walls often 'sweat,' damaging results to wallpaper if this is put on too soon. water is produced as the result of a chemical action between the mortar and the carbon dioxide of the air ($\text{Ca(OH)}_2 + \text{CO}_2 = \text{CaCO}_3 + \text{H}_2\text{O}$). If the air near the wall is already saturated, the additional moisture thus formed is unable to evaporate, and trickles down the wall.

The formation of 'dew' on grass is in some respects similar, but part or all of the deposit may be the result of transpiration from the blades of grass.

Fog and Cloud. Very often the air for some distance above the surface layer is also cooled to dew-point (or, of course, below it), and particles of moisture are formed, giving a *mist* or *fog*.¹ There are always minute particles of dust in the atmosphere, and apparently these serve as the nuclei round which the water vapour is precipitated.

Clouds are usually formed as the result of an upward movement of moisture-laden air. The farther this rises from the earth, the lower becomes its pressure, and consequently it expands (Boyle's law). This expansion produces cooling,² and a cloud is produced—a cloud being a bank of mist or fog condensed at some distance above the earth's surface. The tiny particles of water may run together, forming drops sufficiently large and heavy to fall as *rain*.

To dissipate the fog it would, of course, be necessary to raise the temperature of the air above dew-point. To do this *artificially* requires a large amount of heat, but it is done occasionally to assist airmen landing. Petrol is used as the source of heat, and the process is necessary because of the large quantity required.

Compression of a gas produces *heat*, as when a bicycle pump is worked. With expansion we have the opposite effect. This will be clearer after Chapter 21 has been read.

Questions

1. Describe an experiment to show the difference between unsaturated and a saturated vapour. *Lond.**

2. How would you show that the temperature at which water boils depends on the pressure on the water? Deal with pressures above and below atmospheric pressure, in each case indicate how the pressure is measured and drawing a good diagram. *Camb*

3. Explain what is meant by a saturated vapour.

What is the effect on the pressure of a saturated vapour of (a) decreasing the volume, (b) increasing the temperature?

When the upper fixed point of a thermometer is being determined accurately, care is taken (i) to have the thermometer bulb in contact with steam from boiling water, and not in the water itself, (ii) to correct for the pressure of the atmosphere at the time of the test. Explain as fully as you can the reason for these two precautions. *Ox.*

4. Why, in cold weather, does mist form on the eye-glass of a person entering a warm living-room? *Brist.**

5. Explain why the 'steam' from a railway engine (or a kettle) is invisible near the funnel (or spout). *Camb.**

6. Why does snow disappear in clear sunny weather while the temperature remains below freezing point?

7. Describe an experiment to show that a vapour exerts a pressure and that the vapour pressure at the boiling-point of the liquid is equal to the atmospheric pressure. *C.W.B.**

8. What are the chief differences between evaporation and boiling?

Some water is introduced into the Torricellian vacuum barometer tube of fairly wide bore until a column 5.44 cm. stands above the mercury. The temperature of the laboratory is 20°C . and the atmospheric pressure 753 mm. of mercury. Find the height of the mercury column. (Specific gravity of mercury = 13.6. Saturated vapour pressure of water at 20°C = 17.5 mm.) *J.M.B.*

9. What is meant by the term *dew point*? How would you determine it by experiment?

Under what conditions will hoar frost be formed?

What must be the condition of the atmosphere when the dew point is high, and when it is low in comparison with the temperature? *Dur.*

10. Describe how the following can be demonstrated: (a) that a vapour exerts a pressure, (b) this pressure is a maximum when the vapour is saturated, (c) the maximum vapour pressure is independent of the volume occupied by the vapour.

The maximum vapour pressure at 40°C . of alcohol is 133 mm. of mercury whilst that of ether is 921 mm. State which has the higher boiling point and give reasons for your answer. *J.*

CHAPTER 20

TRANSMISSION OF HEAT

Chapters 18-20 of our earlier book we saw that heat is transmitted by the three processes of conduction, convection and radiation. Here we shall content ourselves with a brief revision of the first two, but shall break up a considerable amount of new ground in Radiation.

We may define **Conduction** as *the process by which heat passes from one portion to another of a body (or of two or more bodies in contact) with an increase in temperature at all intermediate points, but without any change in the position of these points relative to one another.*

An increase in temperature at intermediate points is specified so as to exclude radiation (after we shall be discussing it fully), while the last clause is inserted so as to exclude convection.

The conductivity of various substances in a room may be compared roughly by touch, the best conductors feeling hottest. A much more accurate method is to use heat-conductive paper. Fig. 20/1 represents thin wires of the same dimensions made of copper, brass and iron respectively, laid fan-wise over such paper, and equally heated by a flame placed just under A. We wait until the green band has spread to its maximum extent along each wire. The length of these 'green band' distances then gives the order of conductivities.

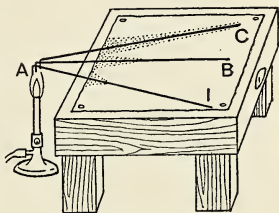


FIG. 20/1

When we try to study the conductivity of liquids, our results are apt to be entirely upset by the fact that convection sets in. To avoid this, we may heat the liquids from the top. Fig. 20/2 illustrates an experiment to show the conductivity of water. The piece of ice is wrapped round

with a little copper wire to make it sink, and it is found that the water at the top may be made to boil while the ice at the bottom is still unmelted.

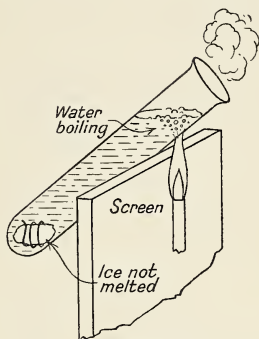


FIG. 20/2

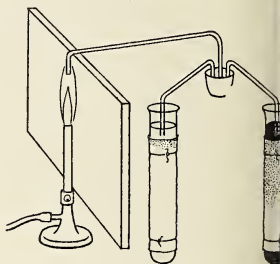


FIG. 20/3

Fig. 20/3 indicates a method of finding which of two liquids (e.g. mercury and water) is the better conductor. The liquids are equally heated, from the top, by means of stout copper wires dipping into a crucible of heated mercury. Heat-sensitive paper is bound to each test-tube.

A good conductor consisting of stout brass gauze forms an essential feature of the Davy lamp (Fig. 20/4). If the explosive gas 'fire-damp' should be present, it can easily pass to the inside of the gauze, where it burns as a blue cap above the ordinary flame. The flame, however, does not pass through the gauze, because the latter conducts the heat away so freely that the explosive mixture on the outside never reaches its 'ignition temperature.'

Bad conductors are very widely used in everyday life. Clothing, blankets, eiderdowns, etc. are all bad conductors, intended to prevent a too rapid loss of heat from the body.

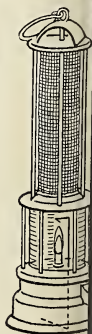


FIG. 20

the leather or rubber sole of a shoe serves the same purpose. It may be wrapped in sacking, flannel, sawdust, etc. to prevent heat from passing into it from the outside air.

Air is itself an exceedingly bad conductor, and so fabrics which contain much air entangled in their substance are very valuable as material for clothing, bedding, etc. Cotton-wool forms a particularly 'warm' packing, say when it is required to prevent the heat of a poultice from escaping too readily. A bird makes use of the bad conductivity of air when it puffs out its feathers in cold weather, and a dog can make its hair stand out in much the same way. A camper prefers two thin blankets to a single thick one, because the former arrangement gives him an extra blanket of air.

All gases, and nearly all liquids, are very bad conductors. Heat is usually transmitted through them not so quickly, however, by *convection*, which we may define as a *process in which heat is conveyed from one part of a fluid¹ to another, by the actual movement of the heated portions of the fluid.*

It is easy to account for this phenomenon. The part of the fluid which is being heated expands. This makes it become less dense than the unheated portions, and so it rises, the place being taken by colder portions of the liquid or gas. The process may be illustrated by the following experiment:—

Support a flask of water by means of a clamp (Fig. 20/5). Introduce a glass tube in the water, reaching to the bottom, and drop a small crystal of potassium permanganate down the tube. Withdraw the latter, first putting your finger over it. Heat the flask by holding the tip of a *small* flame near it. The crystal will form a purple solution, and by means of this we can trace the movement of the water, which is somewhat as indicated by the arrows.

If a liquid underwent no change of volume on heating, convection in it would not be possible. Notice also that *lat. fluere* = to flow. The term includes both liquids and gases.

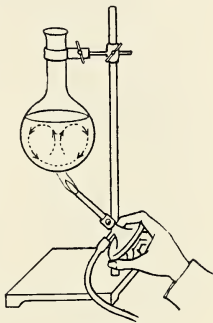


FIG. 20/5

viscosity (stickiness — Lat. *viscum* = birdlime) is a hindrance to convection. That is why jam 'burns' or 'catches' unless it is constantly stirred, and why porridge, though quite cool on the surface, may be very hot underneath.

An application of convection is seen in the hot-water supply of a house (Fig. 20/6). The boiler may be behind the kitchen fire, or at the back of a special stove. The heated water passes into a cylinder, from the top of which a pipe (branching as required) leads to the hot water taps, while an extension, called an expansion pipe, leads to a tank in the loft which is filled from the main. Cold water from this tank flows into the cylinder and from there into the boiler.

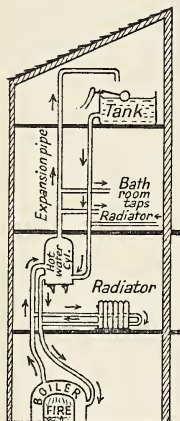


FIG. 20/6

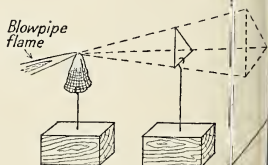


FIG. 20/7

So much for Conduction and Convection. The fact that there must be a third process by which heat is transmitted is illustrated by the passage of heat from the sun to the earth. Since nearly the whole of the intervening distance consists of empty space, it is clear that we are not concerned with conduction or convection. Let us examine some features of *radiation*, the name by which this third process is known.

(1) *It travels in straight lines.* This may be shown by the experiment illustrated in Fig. 20/7.

Support a small tin triangle (sides an inch or so long) in a vertical position. Three or four inches from it support a piece of wire gauze that has been folded into a rough cone, per-

wards. Heat the apex with a small bunsen flame, or (better) with the tip of a blowpipe flame. The idea is to obtain, as nearly as may be, a point source of heat.

While you are doing this a helper pins a sheet of heat-sensitive paper to a drawing-board, and holds it in a vertical position an inch or two from the tin triangle, the latter being between the source of heat and the sensitive paper. The paper soon shows a white triangular patch surrounded by green. The patch corresponds in size and position to the shadow that would have been obtained by using a source of light instead of heat.

Assuming that light travels in straight lines, this experiment shows that the same thing is true of heat.

2) *It travels with great speed.* Suppose it is a hot summer's day, and a cloud passes across the face of the sun. The cooling effect is felt at once. This shows that heat radiated from the sun must take very little time indeed to travel the distance so that separates the cloud from the earth. (Actually the speed is that of light—about 187,000 miles per second.)

3) *Intervening medium not heated.* Even in summer an alpinist at great heights (if not equipped with special clothing) suffers severely from the cold. Heat from the sun is passing through the air, but without heating it. Similarly, the glass in a greenhouse is often much cooler than the air inside—although the heat must have passed through the glass.

We should now have no difficulty in understanding the following definition of **Radiation**. It is a process in which heat is transmitted with great speed, in straight lines, and without heating the medium through which it passes.

At this stage let us summarise our present knowledge.

	Conduction	Convection	Radiation
Medium	Solid, liquid or gas.	Liquid or gas.	No material medium.
Change in medium	Parts maintain same relative positions.	Parts change their relative positions.	
Speed	Rise of temperature. Greatest in metals, but slow at the best.	Rise of temperature. Usually greater than conduction. Impeded by viscosity.	No rise of temperature. Enormous.
Direction	Any direction.	Any direction.	Straight lines.

The Solar Spectrum. We have seen that in several important respects heat seems to be radiated in the same way as light. In neither case is any material medium required and in both cases the heat (or light) travels in a straight line. Most striking resemblance of all, both travel at the same very high speed. We are driven to the conclusion that the subjects of Heat and Light are not fundamentally so different as they appear at first sight. That is indeed the case, and at this point we will recall an experiment which properly belongs to the subject of Light, because we shall thereby get a much more thorough knowledge of the nature of Heat. A fuller account of the experiment is given on p. 260 of our earlier book.

In Fig. 20/8, A is a hole in a shutter through which a narrow beam of sunlight enters an otherwise dark room. At B the

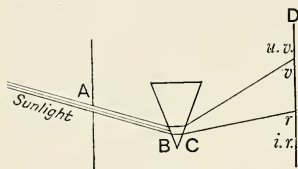


FIG. 20/8

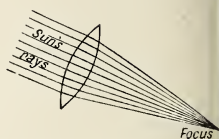
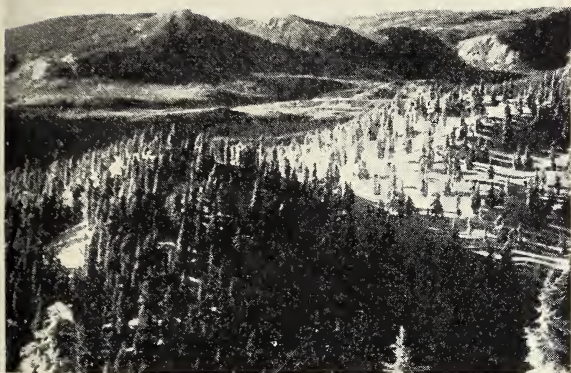


FIG. 20/9

light falls on a prism, where it is *refracted*—i.e. bent out of its course—and it is refracted once more when it leaves the prism at C.

Now, the 'white' light of which the beam consists is really a mixture of light of many different colours, and these are refracted *unequally*—the red least, the violet most, and other colours in between. Thus if the light which emerges from the prism is received on a screen at D, it is found to consist of a band of colour known as a *spectrum*, the end nearest the apex of the prism being red and the other end violet. Such a spectrum would be 'impure'—i.e. the various colours would overlap. By a suitable arrangement of lenses, however, the red rays may all be brought to a *focus* at *r*, the violet rays at *v*, and so on; and we obtain a 'pure spectrum.' The 'suitable arrangement' will be discussed in connection with Light (Chapter 27).



Photograph by permission of Ilford, Ltd.

‘INFRA-RED’ COMPARED WITH ‘ORDINARY’ PHOTOGRAPHY

The different colours correspond to light of different wavelengths. The red waves are about twice as long as the violet but in any case the lengths are so tiny that a special unit called the *micron* ($= 1/1000$ th of a millimetre) is used to express them. The red waves have a length of about 0.0001 micron.

What lies on the other side of the red, the *infra-red* (shown in Fig. 20/8)? Certainly nothing that our eyes can perceive. By the use of the thermopile, however (see below), and using a prism and lenses of rock-salt instead of glass, it can be shown that a large amount of *heat* falls on this part of the spectrum. Here, then, we have a clue to the nature of heat radiation. It is of the same nature as light, differing from it only in the fact that it is produced by longer waves.

If this is the case, we should expect heat to be subject to the same laws of reflection and refraction as light. The fact that heat can be refracted is easily shown by using a convex lens as a 'burning-glass.' It gathers rays of heat to a focus just as easily as it gathers rays of light, and in countries where sunshine is more dependable than in our own, smokers sometimes carry a small pocket lens with them instead of a box of matches (Fig. 20/9).

The fact that heat is reflected according to the same laws as light can also be shown by suitable experiments, though we shall not discuss them here.

It is worth noticing that as infra-red waves are longer than light waves, they are much more effective than light in penetrating a misty atmosphere. That is because particles of dust, water, etc., which are large enough to act as obstacles to the short light waves, are not able to block the longer infra-red ones, and hence the use of the latter in photography (see illustration).

The Thermopile. We must now briefly consider the *thermopile*, to which reference was made above.

If rods of two different metals—say copper and iron—soldered together at the two ends, and a difference of temperature is maintained between the two junctions, a current will flow round the circuit (Fig. 20/10). Now instruments known as *galvanometers* (Chapter 36) are available for detecting extremely small electric currents, and therefore we have

trical means of detecting extremely small differences of temperature.

In practice, the metals antimony and bismuth are found to give a stronger effect than copper and iron. Further, the effect can be multiplied by using a large number of such junctions. The principle is illustrated in Fig. 20/11, where several pairs of junctions are shown. Fig. 20/12a will give an idea of the structure of an actual thermopile. The bars of antimony and bismuth are electrically insulated from one another (except at the



FIG. 20/10

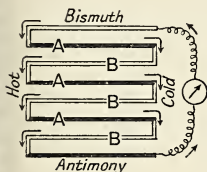


FIG. 20/11

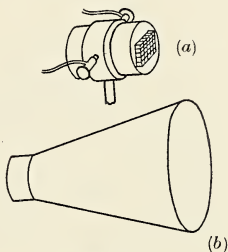


FIG. 20/12

junctions) by strips of mica, and the surfaces intended to receive radiation are blackened. The instrument is often used in connection with a hollow cone, which serves to direct a larger amount of radiation on to the absorbing surface.

Differing absorbing powers. A sheet of white paper reflects most of the light that falls on it, and absorbs only a little. When light falls on black velvet, nearly all of it is absorbed, very little being reflected.

Common experience shows that surfaces must act somewhat differently with regard to heat. The metal curb round a fire-engine often has a black-leaded iron base with a bright brass

The latter is found to be much cooler than the iron, showing that it does not absorb heat so readily. If a motor car has been standing in the sun, the black mud-guard will be found to be much hotter than the chromium-plated

parts close to it. Generally speaking, light-coloured surface absorb less radiation than dark ones, and *polishing* a surface greatly reduces its absorbing power.

Experiments on this subject of absorbing power were carried out as far back as 1805 by the physicist John Leslie. His apparatus consisted of (i) an *air thermoscope*, and (ii) what is now known as a *Leslie's cube*. Both are shown in Fig. 20/13.

The air thermoscope consists essentially of a U-tube containing coloured water (or alcohol). The limbs of the tube terminate in two large bulbs, for which inverted flasks are often used.

The 'cube' is a cubical metal box of about 5 inches side, with a removable cap at the top. For our present purpose (comparison of absorbing powers), the four vertical surfaces would have to be exactly similar.

Suppose we wish to examine the effect of *blackening* a surface. One bulb of the thermoscope would be blackened, the other being left clear. The cube, containing water which has just been boiling, would be placed in such a position near the bulbs that an equal amount of heat would be radiated to each. In this case, the liquid below the blackened bulb would be pushed down, showing that more heat was being absorbed on that side.

Radiating Power. We have just been considering the capacity of different surfaces for *absorbing* heat. We may now think of their capacity for giving it out—their *radiating* power.

Here again we use a Leslie's cube, but this time the four vertical surfaces are differently treated. One may be covered with lampblack, another varnished, another roughened, and the remaining one polished. Instead of the differential air thermoscope, we may use the far more sensitive thermopile. Boiling water is poured into the cube and the cap is replaced. The water is kept boiling by means of a small bunsen flame.

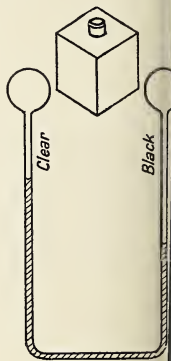


FIG. 20/13

ect heat from the latter being prevented from reaching the thermopile by means of an asbestos screen (not shown). The various faces are now turned in succession to the thermopile, which is connected to a sensitive galvanometer, and the respective deflections are noted.

Now all the faces are obviously at the same temperature, but it is found that they affect the thermopile very differently. The lampblacked surface is found to have the strongest effect. The varnished face comes second, and the rough face third. The polished face produces very little effect.

On comparing the results of such experiments as this with those obtained in a preceding paragraph (for absorbing power), it is found that *good absorbers are also good radiators*.

A simple but rather striking illustration of this statement may be obtained by painting a piece of china, with a dark pattern on a light ground, into a good fire. We now obtain a light pattern on a comparatively dark background.

The good absorber (dark pattern) shows itself to be a

poor radiator, while the poor absorber (represented by the light ground) is seen also to be a poor radiator.

Practical applications. It frequently happens that a surface is required to lose as little heat as possible by radiation. It would be wasteful, for instance, if an electric iron were radiating a great deal of its heat into the air. Hence such a surface is brightly polished. Similarly a bright silver tea-pot loses little heat by radiation, and so does a thermos flask with its silver-coated surfaces.

Hot-water pipes and radiators are usually painted a dull black. Really this does not make very much difference, because (1) a radiator gives off its heat mainly by *convection*, and (2) by radiation.

(2) A *dull* light-coloured surface is not much inferior in radiating power to a black one. A bright surface, however

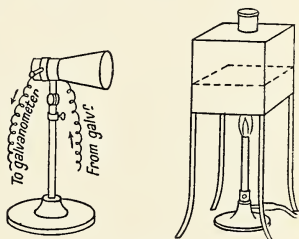


FIG. 20/14

(e.g. when aluminium paint is used), does result in some loss of efficiency.

The vacuum flask. As a conclusion to our studies in conduction, convection and radiation we may consider the vacuum flask, in which devices are adopted for reducing these processes to a minimum. Its essential feature is that it is double walled, with a vacuum between the two walls. A is the point at which the enclosure was sealed off after being exhausted (Fig. 20/15). The outside of the inner wall and the inside of the outer one have been covered with a very thin deposit of silver. The flask is, of course, corked, and for protection is enclosed in a metal case. Chiefly to reduce the risk of breakage from mechanical shock, the flask rests on a cork (or sometimes a spring) placed at the bottom of the metal case.

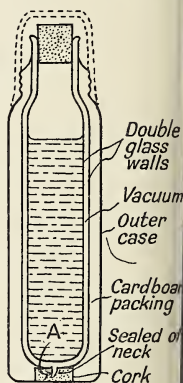


FIG. 20/15

To escape by conduction, heat has to pass to the top of the inner wall—i.e. through a bad conductor (glass) of small cross-section (the glass being thin)—so the rate of loss is very slow.

When the cork is in, the hot liquid is completely cut from contact with the air, so there is no loss by convection.

The inner vessel loses very little heat by radiation because its outer surface is silvered. Of the small amount that does escape, the greater part is reflected by the silvered surface of the outer vessel, very little being absorbed.

When the cork is out, loss of heat by evaporation and convection is considerable, but the remedy is obvious.

Questions

1. A piece of copper gauze is held about an inch above a bunsen burner. The gas is turned on and lit above the gauze when it continues to burn without lighting the stream of gas between the burner and the gauze. Explain. *Oxf.**

2. Describe the Davy safety lamp and explain its mode of action. *Lond.**

3. Describe an experiment to show that copper is a better conductor of heat than iron. *C.W.B.**
4. How would you show that water is a bad conductor of heat? *Lond.**
5. Describe and account for the different sensations experienced touching metal and wooden objects in a refrigerator. *Lond.**
6. Draw a labelled diagram to illustrate the heating of a house by a hot-water system. Explain the process of heat transference through the different parts of the system. *Lond.*
7. Why is a better draught obtained with a high chimney than with a short one? *Camb.**
8. A piece of thin paper is wrapped tightly round a rod one half of which is metal and the other half is wood. Moving the rod back and forth through a small flame, it is found that only one portion of the paper becomes charred. Explain. *Dur.**
9. Explain what is meant by conduction, convection, and radiation.
10. Give a diagram of a thermos flask, and explain how heat interchange between the contents and the surroundings is prevented by its construction.
11. Explain why a thermos flask works more satisfactorily when three-quarters filled with liquid than when it is filled to the brim in a short distance of the cork. *Oxf.*
12. Describe an experiment to show that a dull black surface is a better absorber of heat than a polished surface. *Lond.**
13. Explain what you understand by *radiation of heat*, and describe an experiment to show how the radiating power of a surface depends upon the nature of the surface. *C.W.B.**
14. Write a short account of the various ways in which heat is communicated to a room by a fire.
15. Explain why (a) a flame will not pass through a piece of copper wire until the latter becomes red hot, (b) loosely woven woollen clothing keeps the body warm in comparatively still air but not in strong winds. *Dur.*
16. Compare the properties and nature of radiant heat and visible light, and describe two or three illustrative experiments. *Camb.*

CHAPTER 21

MECHANICAL EQUIVALENT OF HEAT

Historical. Robert Boyle, one of the great English scientists of the 17th century, gave much attention to the subject of heat. He came to the conclusion that heat might be due to the rapid movement of the small particles ('molecules' as we should now call them) of which a body is made up. 'When for example, a smith does hastily hammer a nail,' he said, 'the hammered metal will grow exceedingly hot, and yet the hammer appears not anything to make it so, save the forcible motion of the hammer, which impresses a vehement agitation on the small parts of the iron.'

As the chapter proceeds we shall see that Boyle's view agrees, in broad outline, with those of present-day scientists. In the 18th century, however, men went astray. Noticing that heat would flow from a hot body to a cold one, they came to the conclusion that heat was a material fluid, to which they gave the name *caloric*. If a piece of metal is hammered, the caloric is squeezed out by the force of the blows. It comes to the surface, so to speak, and the metal feels hot.

Some very important observations and experiments bearing on this 'caloric' theory, were described by a Count Rumford in the year 1798. He was director of the arsenal at Munich, and one of his duties was to supervise the boring of cannon. He was very much impressed with the amount of heat produced. According to the theory of that day, caloric had been squeezed out of the brass borings. The effect of drilling was to make them less capable of holding heat, just as the effect of squeezing a sponge is to make it less capable of holding water.

Rumford therefore made an experiment to compare the capacity for heat of the brass borings with the capacity of the undrilled brass, expecting to find that the latter was greater. (The experiment was something like that described on p. 169 for finding specific heat.) He could detect no difference. Apparently heat was not being squeezed out of the brass. Perhaps it was coming from the air?

Accordingly he arranged an experiment in which boring took place under water.¹ After two hours and twenty minutes, the water began to boil.

These and other experiments drove Rumford to the conclusion that heat was not a material substance at all. 'Anything which any *insulated* body, or system of bodies, can continue to furnish *without limitation* cannot possibly be a *material substance*,' he said, 'and it appears to me to be extremely difficult, if not quite impossible, to form any distinct idea of anything capable of being excited and communicated in these experiments, except it be motion.'

In 1799 Davy (the inventor of the lamp) took up the subject. He rubbed together two pieces of ice, with the result that sufficient heat was produced to melt them. He argued that there are only three possibilities:

The heat must have come either (a) from the ice, or (b) from the surrounding air, or (c) from the motion.

Now, if heat has been squeezed out of the ice, the water produced should have a smaller capacity for heat than ice. Actually its capacity for heat is about twice as great. *That explanation will not do.*

To decide whether heat could have come from the air, Davy contrived an experiment in which two pieces of ice were rubbed together in the exhausted receiver of an air-pump. They were still melted. Clearly the heat could not have come from the air.

The only remaining possibility is that heat must be a result of the motion, and Davy gave as his verdict that 'Friction causes vibration of the corpuscles of bodies, and this vibration is heat.'

Thus Boyle, Rumford and Davy had all arrived at what is really the same conclusion, though we should now speak of 'molecules' instead of 'corpuscles.' That conclusion is still accepted.

Further examples. Cases have just been quoted in which kinetic energy has been converted into heat—the kinetic energy of a hammer (Boyle), of a rotating drill or borer

This is open to the criticism that the heat may now be coming from the *water*. Davy settled the point far more satisfactorily by transferring the experiment to a *vacuum*.

(Rumford), of pieces of ice (Davy). We can easily think of other examples. Thus a bullet becomes very hot when it strikes a target; so do the brake-shoes of a car as they press against the revolving drum. As we saw a piece of wood burn, so both wood and saw become hot. The heat produced in pumping up a bicycle tyre is very noticeable; and each of us could, no doubt, give many more examples.

Conversely, heat may be converted into motion. The motion of water through hot-water pipes and radiators is produced by the heat of the boiler fire. Sea-breezes are due to land and water being unequally heated by the sun, while in the steam engine heat from the furnace may cause a heat train to move at sixty miles an hour.

Is there a quantitative relation? Boyle, Rumford and Davy realised that heat was closely connected with motion, but if we think a little we shall see that the motion is not everything. If a cannon-ball strikes a target it will produce much more heat than would a bullet travelling at the same speed. It seems, then, that we must take into account the *mass* of the moving body, as well as its speed. Now, mass and speed are both included in the term energy (p. 45)—or *mechanical energy*, to be more precise. In all the examples we have quoted (the bullet and target for instance) it is mechanical energy that is being converted into heat.

Now, mechanical energy can be accurately measured. Various units are employed, but quite a common one is the foot-pound. Again, heat can be measured, say in British thermal units. The question at once arises, *will a given quantity of mechanical energy always give rise to the same amount of heat?*

The work of Joule. In the 1840's a young Manchester physicist named Joule pondered the question just stated. He devised a number of experiments with a view to answering it, and we will give some account of one of them.

First as to his general principle. Two falling weights were made to turn a paddle which was surrounded by water in a suitable vessel (in Fig. 21/1 C and C' are the weights, and the paddle is inside the vessel E.)

Suppose each weight is 30 lb. and has a fall of 5 ft. Then the work done by a single weight in falling is $30 \times 5 = 150$

b. The work done by the *two* weights would, of course, 300 ft.lb., and if they fall not just once, but twenty times, the work done would be $300 \times 20 = 6000$ ft.lb.

So much for the work done. Now for the heat produced. Suppose the water in the container weighs 8 lb., and the rise in temperature produced by the action of the paddle is 1° Fahrenheit. Then the amount of heat produced = 8 British thermal units. To produce this, the work that had to be done = 6000 ft.lb.

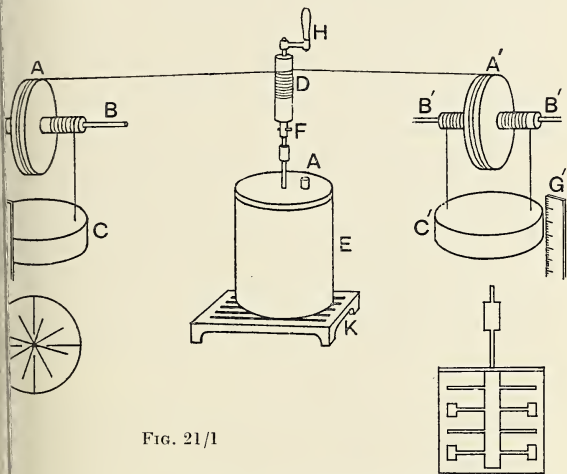


FIG. 21/1

to produce *one* British thermal unit the work required $6000 \div 8 = 750$ ft.lb.

Let us now consider Joule's apparatus and experiment in more detail.

The two pulley-wheels AA' have their axles BB, B'B' mounted in a special way (not shown in the figure), so as to rotate with a minimum of friction. The falling weights CC' are made of lead, and weigh about 29 lb. each. It will be seen from the figure that as the weights fall they cause the shaft D to turn round, and this in turn causes rotation of the paddle which is inside the copper vessel E.

The paddle is shown separately on the left and right (plan and elevation respectively). It consists of eight blades making angles of 45° with each other, and turns through four 'vanes' fixed to the sides of the vessel. These vanes have holes cut in them corresponding in shape to the blades and the latter turn through them much as a key turns in the wards of a lock. The vanes serve to prevent the water from *swirling* as the paddle turns. In other words, the paddle meets with great frictional resistance as it turns through the water, and the friction will give rise to heat.

When the weights have fallen, the pin F is taken out to disconnect the spindle from the paddle. By turning H the weights are wound up. F is then replaced, the weights are allowed to fall again, and so on. The distance they fall can be read by means of scales G, G'.

The temperature of the water in the vessel was recorded by a thermometer which passed through the aperture. This thermometer was very sensitive, and Joule was able to record changes of as little as $1/200$ th of a degree Fahrenheit.

To reduce loss of heat by conduction, the vessel E stood on a wooden grid K. For the same purpose a boxwood insertion was placed in the central axle.

Let us now look back to the calculation on p. 209. There are two very obvious corrections that must be applied if we expect to obtain a reasonably correct result.

- (i) We must allow for the water equivalent of the metal vessel E—really a calorimeter—and of the paddle wheel.
- (ii) One of Joule's experiments would last about half an hour, and during this time the calorimeter would be losing heat by radiation (or gaining heat, if the temperature of the enclosure was higher than that of the calorimeter). He corrected for this by allowing the calorimeter to remain for another half-hour (or whatever the time was), and noting the change of temperature. From the data thus obtained he was able to calculate what the rise of temperature in the actual experiment *would* have been if there had been no loss (or gain) by radiation.

It gives us some idea of Joule's extreme carefulness when we notice that he took the trouble to look out a 'spacious cellar' for his experiments, so as to secure uniformity of temperature, and that he placed a large wooden screen between himself and his apparatus, so that the temperature of the latter should not be raised by radiation from his body.

He made a number of corrections not mentioned in the general account given above. For instance, he allowed for the weight of the string to which his lead weights were attached; and he allowed for the fact that his falling weights did not give up *quite* their energy to the paddle. He found that on reaching the floor they had a velocity of 2.42 in. per sec., and thus a little kinetic energy was given up to the floor in the form of heat.

Joule devised other experiments besides the one we have described, and obtained scores of results. His final conclusion was that what we now call 1 British Thermal Unit is equivalent to 772 ft.lb. More recent results, notably by H. A. Rowland, give the value 778. We express this by saying that *the mechanical equivalent of heat is equal to 778 ft.lb. per British Thermal Unit.*

In honour of Joule, the mechanical equivalent is usually noted by J—i.e. $J = 778$ ft.lb. per British Thermal Unit. To raise the temperature of 1 lb. of water by 1° C. requires, of course, $9/5$ ths as much heat as to raise it 1° F.—i.e. the 'lb.-degree C.' unit is $9/5$ ths as large as the lb.-deg. F. unit (B.Th.U.).

\therefore the value of J in lb.-degree C. units is $778 \times \frac{9}{5} = 1400$.

Let us find the value of J in the 'c.g.s.' system of units, given that 1 lb. = 454 gm. and 1 ft. = 30.5 cm.

778 ft.lb = work done in raising 778 lb. through 1 ft.

= work done in raising 778×454 gm. through 30.5 cm.

= work done in raising $778 \times 454 \times 30.5$ gm. through 1 cm.

= 10,800,000 centimetre-grams (approx.)

= $10,800,000 \times 981$ ergs (p. 41)

= 1060×10^7 ergs (approx.).

Again, 1 B.Th.U.

= heat required to raise temperature of 1 lb. of water by 1° F.

= heat required to raise temperature of 454 gm. of water by $5/9^{\circ}$ C.

= heat required to raise temperature of $454 \times \frac{5}{9}$ (= 252) gm. of water by 1° C.

= 252 calories.

Thus instead of saying

$J = 778 \text{ ft.lb. per British Thermal Unit}$ we may say

$J = 1060 \times 10^7 \text{ ergs per 252 cal.}$

$= \frac{1060}{252} \times 10^7 \text{ ergs per cal.}$

$= 4.2 \times 10^7 \text{ ergs per calorie.}$

Summarising the various ways in which the value of J may be expressed, we have:—

$J = 778 \text{ ft.lb. per British Thermal Unit}$

$= 1400 \text{ ft.lb. per lb.-deg. C.}$

$= 4.2 \times 10^7 \text{ ergs per cal.}$

$= 4.2 \text{ Joules per cal.}$

(a Joule being defined as 10^7 ergs).

Why Joule's work is outstanding. One could name many men in the 19th century who carried out numerous experiments with the utmost patience and accuracy. But Joule's work is regarded as quite specially important. Why?

There are in science certain great *generalisations*—truths which may be simple, but which are of very wide application. They apply to everything in the world and sometimes to everything in the universe. One of these generalisations is Newton's law of gravitation. We have to take it into account whether we are trying to explain the fall of an apple or the motion of a planet. Another is the theory of evolution put forward by Charles Darwin, which has thrown light on a thousand problems applying to all life on the earth, problems which would otherwise have been utterly baffling.

Still another is the principle of the 'Conservation of Energy'—the principle that energy can be neither created nor destroyed, though it may be converted from one form into another. Apparently it applies to the whole universe. As we have seen, Joule showed that there was a very exact relationship between mechanical energy and heat. By other experiments he showed that there was a similar relationship between chemical energy and heat, and between electrical energy and heat. Thus his work gave a firm basis for the Principle of Conservation of Energy, and it is because that principle is one of the great generalisations of science that the work of Joule is regarded as being of quite outstanding importance.

Laboratory Methods. We shall describe two laboratory methods for finding J .

- (i) The falling lead-shot method.
- (ii) The rotating drum method.

(i) *The falling lead-shot method.*

A cardboard tube AB about a yard long and 3 or 4 in. wide corked at both ends, and about 4 lb. of lead-shot is introduced (C). The temperature of the shot is taken with a sensitive thermometer. The tube is then completely inverted for a definite number of times, say 100, and the temperature is then taken again. The average distance fallen by each lead-shot must so be measured (it is represented by h in Fig. 21/2). We will take it that the specific heat of lead is known ($= 0.031$). For our unit of heat we will take the pound degree-Centigrade unit, i.e. the amount of heat required to raise the temperature of a pound of water by 1°C .

Suppose that in a given experiment $h = 3.25 \text{ ft}$.

Initial temperature of lead-shot $= 14.72^\circ \text{C}$.

Final " " " " $= 18.89^\circ \text{C}$.

No. of falls $= 100$.

The weight of the lead-shot need not be known.

Let it be $w \text{ lb}$. Then—

Work done by $w \text{ lb}$. of lead, falling 3.25 ft . $= 3.25w \text{ ft.lb}$.

\therefore work done in 100 falls $= 3.25w \times 100 \text{ ft.lb}$.
 $= 325w \text{ ft.lb}$.

Heat produced = heat required to raise temperature of $w \text{ lb}$. of lead by $(18.89 - 14.72)^\circ \text{C}$, i.e. by 4.17°C . This amount is $0.031 \times w \times 4.17 \text{ units}$.

$\therefore 0.031 \times w \times 4.17 \text{ units of heat are produced by } 325w \text{ ft.lb}$.

$\therefore 1 \text{ unit of heat is produced by } \frac{325w}{0.031 \times w \times 4.17}$
 $= 2514 \text{ ft.lb}$.

As the correct answer is 1400, this result is absurdly high, the important source of error being that, at every inversion, each of the lead shot falls before the tube has reached the vertical position. However, the experiment helps to fix the idea of a quantitative relation between work and heat, and has the advantage of great simplicity.

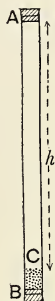


FIG. 21/2

The following method gives much more accurate results:—

(ii) *The rotating-drum method.*

The apparatus consists of a brass drum which can be turned by means of a handle. As it turns, it presses against a brake consisting of a silk ribbon, kept tight by a known weight attached to its lower end and a spring balance pulling at its upper end. The weight (say w gm. wt.) is balanced by the frictional force (f gm. wt.) between drum and ribbon, *plus* the pull (s gm. wt.) of the spring balance; i.e. $w = f + s$, and so $f = w - s$.

As silk is a bad conductor, not much of the heat produced on its inner surface is lost by passing outwards—nearly all of it passes inwards through the drum (made of good conducting material) to a known weight of water contained within. The temperature of this water is taken by a bent thermometer, which passes through a hole in the drum.

To carry out an experiment, we put a measured quantity of water into the drum by means of a burette, and read the thermometer. (This has a very open scale reading to 0.01° C.). We then turn the handle several hundred times at a speed sufficient to keep the weight floating and to maintain the reading of the spring balance about the centre of the scale. Suppose this reading is 160 gm. and the hanging weight is 700 gm., then as already explained to turn the drum it is necessary to overcome a frictional force of $700 - 160$ or 540 gm. In *one* turn the work done against this force of friction is $540 \times C$ centimetre-grams, where

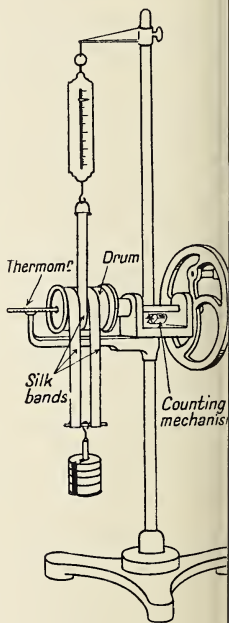


FIG. 21/3

the circumference of the drum measured in centimetres. To find the total work done we must multiply by the number of turns recorded by a counter included in the instrument.

Of the heat produced, some goes into the drum as well as into the water, so we shall have to find the water equivalent of the drum.

We shall now be able to follow the working out of results such might be obtained in an actual experiment.

Mass of brass drum (given)	. . .	272 gm.
Circumference of drum	. . .	23.5 cm.
Water added = 50 c.c.	. . .	50 gm.
Specific heat of brass	. . .	0.092
Initial temperature of water	. . .	15.26° C.
Final	„ „	16.83° C.
∴ Rise of temperature	. . .	1.57° C.
No. of revolutions	. . .	400
Lower weight	. . .	700 gm.
Reading of spring balance (average)	. . .	160 „

Work done = $(700 - 160) \times 23.5 \times 400 = 5.08 \times 10^6$ cm.-gm.

Heat produced = $(50 + 272 \times 0.092) \times 1.57 = 118.0$ cal.

∴ 118 cal. = 5.08×10^6 cm.-gm.

∴ 1 cal. = 4.31×10^4 cm.-gm.

Mechanical equivalent of Heat = 4.31×10^4 cm.-gm. per cal.

By working through calculations such as the following we shall improve our grip of the fact that a given amount of mechanical energy is equivalent to a definite amount of heat.

Example 1. By how much should the temperature at the bottom of Niagara Falls (height 167 ft.) exceed that at the top? ($J = 778$ ft.lb. or B.Th.U.)

The rise of temperature will be the same whatever weight of water we consider,¹ so for simplicity let us consider 1 lb.

Work done by 1 lb. of water falling 167 ft. = 167 ft.lb.

778 ft.lb. produce 1 B.Th.U.

∴ 167 „ „ $\frac{167}{778}$ „

Now 1 B.Th.U. would cause temp. of 1 lb. of water to rise 1° F.

∴ $\frac{167}{778}$ B.Th.U. make it rise $\frac{167}{778}$ ° F. = 0.21° F. or 0.12° C.

(N.B. Water falling over rocks would lose much of its kinetic energy en route, by friction against the rocks, which would receive a large part of the heat produced. Cf. next example.)

¹ Cf. the lead shot experiment (p. 213) in which the w 's cancel.

Example 2. A brick weighing 9 lb. slides 12 ft. down a rough plank inclined at 30° to the horizontal, and acquires a velocity of 10 ft. per sec. How much heat is produced?

The vertical height through which the brick has fallen is 6 ft. (easily shown by geometry).

\therefore work done by gravity $= 9 \times 6 = 54$ ft.lb.

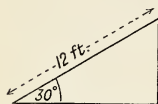


FIG. 21/4

At the bottom of the plane, the kinetic energy actually possessed by the brick ($\frac{1}{2}mv^2$) is $\frac{1}{2} \times 9 \times 10^2 = 450$ ft.pdls.

$$= 450 \div 32 \text{ ft.lb.} = 14.1 \text{ ft.lb.}$$

The difference $54 - 14.1$ or 39.9 ft.lb. represents the energy which has disappeared as heat (the result of friction between the brick and board).

Now 1400 ft.lb. $= 1$ unit of heat (lb. deg. C.)

$$\therefore 39.9 \text{ ,,} = \frac{39.9}{1400} \text{ or } 0.0285 \text{ unit.}$$

Example 3. A stationary engine burning $37\frac{1}{2}$ lb. of coal per hour is able to pump water to a height of 490 ft. at the rate of $62\frac{1}{2}$ galls. per minute. Find the efficiency of the engine (1 lb. of the coal gives 7000 lb.-deg. C. units; 1 lb.-deg. C. $= 1400$ ft.lb.; 1 gal. $= 10$ lb.)

(a) Heat produced per minute.

Coal burnt per hour $= 37\frac{1}{2}$ lb.

$$\therefore \text{heat produced per hour} = 37\frac{1}{2} \times 7000 = 262,500 \text{ units}$$

$$\therefore \text{,, ,, ,, min.} = 262,500 \div 60 \\ = 4375 \text{ units.}$$

(b) Work equivalent to this heat.

$$1 \text{ unit} = 1400 \text{ ft.lb.}$$

$$\therefore 4375 \text{ units} = 4375 \times 1400 \text{ ft.lb.}$$

(c) Work actually done per minute.

$62\frac{1}{2}$ galls. ($= 625$ lb.) is raised 490 ft.

$$\therefore \text{work done} = 625 \times 490 \text{ ft.lb.}$$

(d) Efficiency. This is the percentage that (c) is of (b).

$$\therefore \text{efficiency} = \frac{625 \times 490}{4375 \times 1400} \times 100$$

which works out to 5%.

Joule. James Prescott Joule was so outstanding a figure in the history of 19th-century Physics that we ought to know something of his life.¹

He was born at Salford near Manchester on Christmas Eve 1818. As a boy he was considered rather too delicate for ordinary school life, and so he was educated at home until the age of sixteen. Soon afterwards he studied geometry and chemistry under John Dalton. Maybe he caught something of Dalton's independent scientific outlook. Certainly it is interesting to note that one of the century's greatest physicists should have studied under one of its greatest chemists.

In his early twenties he was much interested in the 'electromagnetic engine' (i.e. the electric motor), and he tried to measure its power by the rate at which the magnets contained it would lift a weight. Later he realised that the amount of work he would get out of his 'engine' was closely related to the consumption of zinc in the battery; in other words, he saw that mechanical energy might be the equivalent of a definite amount of chemical energy. Before he was twenty-five he had a very clear grip of the problem that was to dominate his life—the quantitative relationship between mechanical energy and heat. The 'paddle-wheel' experiment already described is only one of a number of methods which he devised in his attempt to solve the problem. It was carried out in 1845.

The idea that a given amount of mechanical work could be converted into a definite quantity of heat was a new one, and, like most new ideas, it had a somewhat chilly reception. This is very clear from what happened in 1847. Joule was to read a paper on the subject to a section of the British Association which in that year met at Oxford. Time was rather short and (writes Joule) 'the chairman suggested that I should not read my paper but confine myself to a short verbal description of my experiments.'

A sure sign that the chairman did not consider the paper to be of much importance !

For most of the particulars in this paragraph the writer acknowledges his debt to 'Joule and the Study of Energy,' by A. Wood (published by Bell).



George Foster A.R.A. circa 1863

H. Maynard

2^d That the quantity of heat capable of increasing the temperature of a lb of water (weighed in vacuo, and taken at between 55 and 60°) one degree Fahrenheit, requires for its evolution the expenditure of a mechanical force represented by the pressure of 772 lbs through the space of one foot.

J. P. Joule

By courtesy of the Manchester Literary and Philosophical

JAMES PRESCOTT JOULE

'This I endeavoured to do,' Joule continued, 'and discussion not being invited, the communication would have passed without comment if a young man had not risen in the Section, and by his intelligent observations created a lively interest in the new theory. The young man was William Thomson' (later the famous Lord Kelvin).

It is interesting to read Lord Kelvin's account of the same incident. 'I can never forget the British Association at Oxford in 1847,' he says, 'when in one of the sections I heard a paper read by a very unassuming young man, who betrayed no consciousness in his manner that he had a great idea to unfold. I was tremendously struck with the paper. I at first thought it could not be true, because it was different from Carnot's theory, and immediately after the reading of the paper I had a few words with the author, James Joule, which was the beginning of our forty years' acquaintance and friendship.'

'Very unassuming,' says Kelvin, and like most truly great men, Joule was essentially humble. Writing to his brother in 1887 he writes, 'I believe I have done two or three little things, but nothing to make a fuss about.' He died in 1889, and in Westminster Abbey, near the burial places of Lord Kelvin and Sir Isaac Newton, there is a tablet to his memory.

Questions

A

1. Give *three* everyday examples of the conversion of work to heat.
2. Give *two* examples of the conversion of heat into work.
3. How many British thermal units are equivalent to 7780 lb.?
4. (a) An engine is working at 10 h.p. How much work is done in one minute? (Answer in factors.)
(b) If this work were all made to pass as heat into a gallon of water (= 10 lb.) by how much would the temperature be raised?

B

1. Describe a non-electrical method of finding the relation between work and heat, and define the units of work and heat employed.

A steam plant, burning 0.5 lb. of coal per minute, supplies steam to a steam engine. If the overall efficiency of the system is 25%, determine the horse-power generated by the steam engine. The calorific value of the coal = 14,960 B.Th.U. per lb., the mechanical equivalent of heat = 780 ft.lb.wt. per B.Th.U., and 1 h.p. = 33,000 ft.lb.wt. per min. *C.W.B.*

2. A man walks up a flight of stairs 10 metres high. If half the energy which he expends is converted into heat, how much would you expect his temperature to increase if there were no natural mechanism for keeping it constant?

(1 cal. = 4.2 joules. Assume that the heat capacity of the human body is the same as that of an equal mass of water. $g = 981$ cm. per sec.²) *Camb.**

3. The *mechanical equivalent of heat* is 778 ft.lb. per British Thermal Unit. Explain this statement carefully, defining the terms printed in italics.

A load of 74 lb. is raised 3 ft. vertically by means of a machine whose efficiency is 22.2%. Assuming that the portion of the work done by the effort that is apparently wasted eventually reappears as heat, calculate the number of B.Th.U. produced. *Oxf.*

4. If all the heat which is given out by 1 kg. of steam in condensing to water at 100° C. is converted into work, how many joules will be produced? If the work is done in 1 minute, what is the average power produced? (Latent heat of steam = 540 cal. per gm.; mechanical equivalent of heat = 4.2 joules per cal.) *Camb.*

5. Water flows over a waterfall 30 metres high, and at the foot of the fall it drops into a pool where it may be regarded as stationary. What is the difference in temperature between the water at the top and that at the bottom of the fall?

(1 cal. = 4.2×10^7 ergs, $g = 981$ cm. per sec. per sec.) *O.C.**

6. 100 gm. of lead shot is allowed to fall 50 times through a tube 1 metre in length and is then dropped into 20 gm. of water the temperature of which is raised 0.59° C. Assuming that the mechanical energy is converted into heat, calculate the mechanical equivalent of heat in ergs per calorie. *O.C.**

CHAPTER 22

LIGHT—PROPAGATION AND REFLECTION

A. Propagation

WE have many reminders in everyday life that light travels in straight lines. We notice the straight beam from the headlight of a car, from a cinema projector, or the humble electric torch. A burly individual keeping his hat on at the pictures can very effectively obstruct the view of the unlucky person sitting behind—the light is unable to travel *round* him; and an umpire giving an l.b.w. decision at cricket unconsciously perhaps, assuming the principle in question.

A very simple experiment illustrating the point was described in our earlier book. We had three pieces of card-board, A, B, C, in which were bored three small holes, P, Q

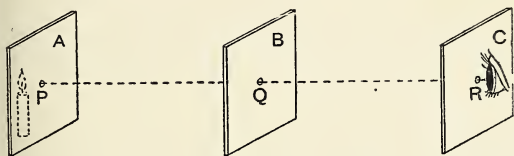


FIG. 22/1

and R, respectively, brought into line by stretching a thread through them. A lighted candle was then placed to the left of P, and observed by an eye placed to the right of R. The flame is visible, but becomes invisible if Q is displaced slightly to one side. This shows that light can travel along the path PQR only when that path is a straight line.

We also studied the *pinhole camera*. The image is inverted because of the crossing, at the pinhole, of the rays of light which form it. If A (Fig. 22/2) is enlarged, more light enters

and a brighter image is obtained. The image, however, is not so sharply outlined. That is because the one large hole may be regarded as equivalent to several small holes very close together. Each of these would produce an image, but the various images would be in slightly different positions.

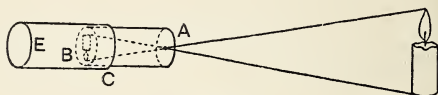


FIG. 22/2

(cf. Fig. 22/3, where two such small holes are represented). Thus the outline is confused.

If the distance from candle to pinhole is increased the image becomes smaller. This result easily follows from similar triangles, for if i = length of image and o = length of object while x = distance of image from pinhole and y = distance of object (Fig. 22/4), then $\frac{i}{x} = \frac{o}{y}$, or $i = x \times \frac{o}{y}$. Clearly if we increase y , we reduce i .

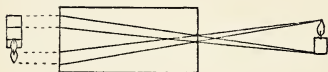


FIG. 22/3



FIG. 22/4

The *shape* of the hole makes no difference so long as it is small. If the hole is circular, the image will be made up of a vast number of circular patches—one circular patch for each point on the object. If it is triangular, the image will be made up of a vast number of triangular patches. It is as though a painter filled in an outline with a multitude of tiny circular patches of paint. He would get the same result if the patches were of triangular or any other shape—so long as they were very small.

Lastly, in our earlier work we studied eclipses of the sun and moon. Fig. 22/5 represents an eclipse of the sun caused by the moon B coming between the sun A and the earth.



ACTION OF THE PINHOLE CAMERA

From a drawing in Guilleman's 'Forces of Nature'

posed to be situated at M or N. The earth's distance from the sun is not always the same).

The shadow of the moon is seen reaching to C. If the moon is near enough for this shadow to touch it (Fig. 22/6), total eclipse will be observed over a small circular patch (over more than 190 miles in diameter). Chiefly owing to

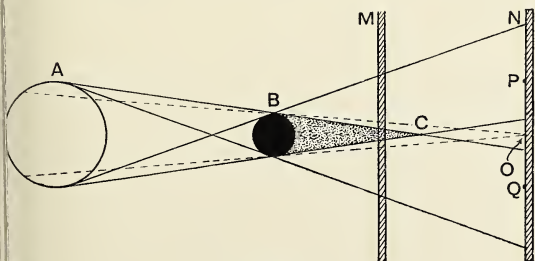


FIG. 22/5

the moon's motion, this circular patch moves over the earth's surface, so that the eclipse is observed first at one part and then another of a track which may be 190 miles wide.

If the earth is beyond the point C, a total eclipse will not be visible at any part of it. At a point such as O, however (Fig. 22/5), between the two 'tangent' lines, only the outer rim of the sun will be observed—i.e. there will be an *annular* eclipse. At P only the top part of the sun will be visible, and at Q only the bottom part—i.e. at these points partial eclipses will be observed.



FIG. 22/6

An eclipse of the moon is produced when the earth comes between the sun and the moon, with the result that its shadow is cast on the moon. Fig. 22/5 will still serve as our illustration, if we now understand that A is the sun and B the earth while M represents the position of the moon. In this case however, the moon is always well on the 'umbra' side of the earth. In the course of its journey round the earth, it happens sometimes that the moon moves into the space represented by the shading in Fig. 22/5. We soon have the appearance

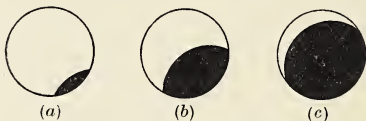


FIG. 22/7

shown in Fig. 22/7 (a), then, later, (b) and (c), and finally the moon is entirely covered by the earth's shadow. Actually the shadow is not completely black. That is because, though the rays of the sun cannot reach the moon by a *straight* path, some of them manage to reach it by a *curved* one (owing to refraction by the earth's atmosphere). The only 'successful' rays are those of greatest wave-length, the red and yellow ones; hence the moon appears as a copper-coloured disc.

From an examination of Fig. 22/8 it should be clear that

an eclipse of the sun can occur only when the moon is in the 'new moon' position, and an eclipse of the moon only when it is in the 'full moon' position.

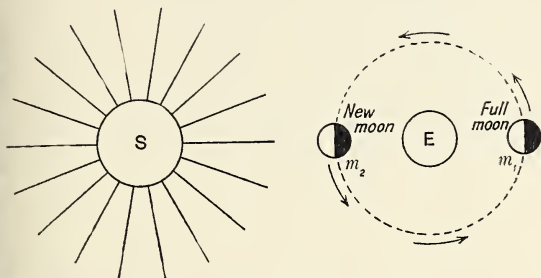


FIG. 22/8

Velocity of Light. One often meets with references to the velocity of light. This has been determined by a number of methods of which we shall indicate one or two.

The first dates back to the year 1676, and is not at all difficult to understand. The planet Jupiter (J_1 or J_2 in Fig. 22/9) possesses a number of 'moons' or satellites, of which, however, we are concerned only with one, M_1 or M_2 . At regular intervals this satellite passes into the shadow cast by Jupiter, and is then eclipsed. The young Danish astronomer Olaf Rømer noted the exact time that elapsed between one such eclipse and the next. Actually it was 42 hr. 28 min. 36 sec.), and it was then

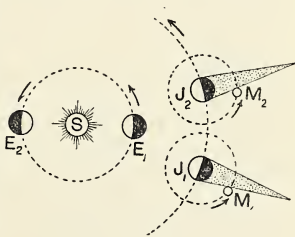


FIG. 22/9

a simple matter to forecast the exact time at which eclipses would occur for months ahead—or at least so Rømer thought. Six months later, however (Jupiter being at J_2 and the Earth at E_2), he found that instead of Jupiter's satellite being

eclipsed at the calculated moment, it was 'behind time' 16 min. 36 sec. ! He was very surprised at first, but after little hard thinking he hit upon the reason. His first observations were made when the earth was, say, at E_1 . Six months later the eclipse would have occurred at the exact time calculated *if the earth had remained at E_1* . But in this six months the earth had moved to E_2 , right at the other end of the diameter of its orbit, a distance of about 186 million miles. The eclipse was 16 min. 36 sec. late because light had taken that time to make its long journey of 186 million miles. We can easily calculate the velocity, for if light takes

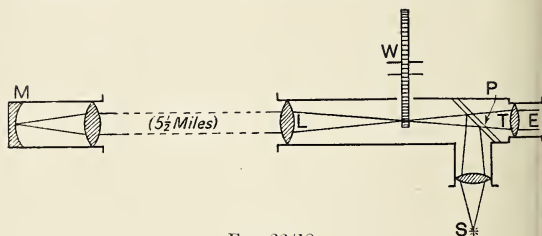


FIG. 22/10

16 min. 36 sec. (= 996 sec.) to travel 186 million miles, the distance travelled per second is $186,000,000 \div 996$ or near 187,000 miles.

In 1849 Fizeau employed an ingenious method in which he made use of a toothed wheel revolving at high speed. The apparatus extended from the village of Suresnes, near Cloud (outside Paris) to Montmartre, a suburb of Paris, a distance of 8.633 kilometres or about $5\frac{1}{2}$ miles.

At Suresnes is S, a source of light. The light was made to strike on the sheet of plate glass P, and then to pass through the space between two teeth of the wheel W. After passing through the lens L it goes on its $5\frac{1}{2}$ -mile journey to the mirror M (at Montmartre), from which it is reflected back along its original path. Some of the reflected light passes through the plate glass into a lens T, and reaches an observer placed at E. To the observer thus placed, the source of light S appeared like a star.

Now, by making the wheel revolve at a certain speed, it was found that S ceased to be observed. We can see what happened. The *space* through which the light had passed on its outward journey is now occupied by a *tooth*, owing to the movement of the wheel. By doubling the speed the light was made to reappear—it was coming back through the *next* space. At treble the speed the light was eclipsed once more, and so on.

The wheel contained 720 teeth and 720 spaces, and the first eclipse occurred when it was making 12.6 revs. per sec. We must calculate the time for it to move forward a distance 'one tooth.'

One tooth occupies $\frac{1}{2 \times 720}$ or $\frac{1}{1440}$ th part of the circumference.

Time required for 1 revolution = $\frac{1}{12.6}$ sec.

time required to move forward 'one tooth' = $\frac{1}{12.6 \times 1440}$ sec.

In this time light has travelled *twice* the distance from W to M, i.e. 2×8.633 or 17.266 km.

no. of km. per sec. = $17.266 \times 12.6 \times 1440$
 = 313,000 km. per sec.
 = 195,000 miles per sec.

(Fizeau's result was about 5% too high.)

A year later—i.e. in 1850—Foucault made experiments in which he used a revolving mirror instead of a revolving wheel, but we will not describe his method here. Since then the problem has been attacked by a number of different workers using very ingenious methods. Some of the best results were obtained by the great American physicist A. A. Michelson (1852–1931), who found the velocity to be 186,271 miles per second. For most practical purposes it is sufficient to remember the number as 186,000.

At this enormous speed, light could travel a distance equal to the earth's circumference in rather less than one-seventh of a second. The sun is so distant, however, that light travelling from it takes about $8\frac{1}{2}$ min. to reach us. If we

could flash a light signal to Jupiter and have it instantly returned, we should receive our reply in about $1\frac{1}{2}$ hr.

Once outside the solar system, these distances and time are enormously increased. The *nearest* fixed star, Alpha Centauri, is at such a distance that light from it takes 4 *years* to reach us, while for the Pole Star the time is 55 years. In fact, if an observer on the Pole Star could witness even on the earth by means of a telescope, he would in 1952 be witnessing the Queen Victoria Diamond Jubilee celebration of 1897.

To express such distances the astronomer often uses a unit known as the *light-year*, this being the distance that light would travel in one year (it comes to about 6 million million miles). Thus the distance of the Pole Star is 55 light-years. That of some of the spiral nebulae is about 100 million light-years—but confronted with such magnitudes our imaginations are helpless.

B. . Reflection

Reflection in Nature. There are a few objects, such as the sun, a candle flame, a fire, etc., which are *self-luminous*; we see them by their own light. The vast majority, however, are seen by the light which has first fallen on them from a self-luminous body, and has then been thrown off or *reflected*. It is interesting to think for a few minutes of the changes we should notice if there were no such thing as the reflection of light. We should see the sun and the stars, but the sky between them would be inky black, and we should see nothing whatever of the moon or the planets which shine by reflected light. It would be useless to go to school. If lights were switched on we should see them, but we should be unable to see the teacher, or the blackboard, or our books; and we should be constantly colliding with other people because we should be unable to see them either. You can trace out a few other consequences for yourself. Life could certainly not continue for very long in a world where no light was reflected.

Laws of Reflection. In our earlier book we considered the *laws* of reflection, which are so important that we must recapitulate them here. They are—

1. *The angle of reflection is equal to the angle of incidence.*
2. *The reflected ray lies in the plane containing the incident ray and the normal.*

These laws can be proved by means of the ray-box (*Junior Physics*, pp. 211–12). A proof in which we use pins instead of a ray-box is perhaps less interesting, but is in fact more exact.

Experiment 1. Place a rectangular strip of mirror glass AB, $6\text{ in.} \times 1\text{ in.}$, on its edge on the bench, lying on a sheet of paper. (It is convenient to have the mirror glued to a block of wood CD which keeps it from falling over.) Stick two pins P, Q, upright in the paper, and place two other pins R, S in such a position that they 'cover' P' and Q' (P' and Q' being the images of P and Q respectively).

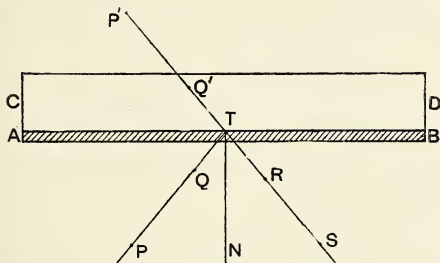


FIG. 22/11

Now PQ represents a ray of light striking the mirror. After reflection, we see the pins in the direction P' Q' (this direction being marked by the pins R and S). This means that the ray of light which reached the mirror by the path PQ has left it by the path RS. With a pencil, draw the outline of the mirror, and then remove it.

Join PQ and SR and produce them to meet at T. It will be found that T is on, or very near, the back of the mirror, from which reflection has taken place. From T draw the normal TN, and carefully measure PTN and STN, the angles of incidence and reflection respectively. Repeat the experiment for various angles of incidence.

Our experiments show clearly that the angles of incidence and reflection are equal. Further, consider the points P, Q, at which the pins enter the paper. PQ represents an

incident ray which grazes the paper—i.e. the ray is *in the plane of the paper*. So is the normal TN. Now P'Q', the bottom ends of the 'image' pins, are also seen in this same plane. Thus we may say that incident ray, normal and reflected ray all lie in the same plane.

Image behind mirror. When a man stands in front of a plane mirror, his image *seems* to be as far behind the mirror as he himself is in front of it. It can easily be shown by experiment that this is actually the case. In this experiment we make use of what is called the *method of parallax*, which you can quickly illustrate for yourself. Hold your fountain pen A about a foot in front of your face, and your pencil about 6 in. beyond the pen. Now move your head from side to side. The pen seems to move faster than the pencil. Interchange these positions. Now, of course, the pencil seems to move faster than the pen. Next make A touch B and move your head once more. This time A and B keep together. Let us put the matter the other way round. When A and B keep together as you move your head, they are in the same position.

All this, of course, is extremely simple, but the principle is often useful, and will enable us to 'fix' the position of an image behind a mirror.

Experiment 2. Place the rectangular mirror on its edge as in Expt. 1, and stick a pin A in the bench, vertically, about 6 in. in front of it. Now place a second pin behind the mirror so that you can see its upper part, and so that it seems to coincide with the image of A. Move your head to see if there is any apparent separation between the real pin and the image. If so, alter the position of the second pin until, on moving your head, pin and image keep together. The second pin is now in the position occupied by the image of the first one.

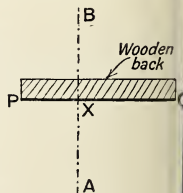


FIG. 22/12

Suppose A is the first pin and B is the position of its image which we have now found. Draw a line PQ to show the position of the (back) reflecting surface of the mirror. Join AB cutting PQ in X. We find that $XA = XB$, and that

perpendicular to PQ. Thus if an object is placed in front of plane mirror, the line joining it to its image will be at right angles to the mirror; and the image will be as far behind the mirror as the object is in front.

It is quite easy to prove geometrically that this result follows from the laws of reflection.

Let A be the pin, PQ the reflecting surface (Fig. 22/13). AC any ray of light proceeding from A and striking the mirror. Draw the normal CN, and construct the reflected ray CR making the angle $\text{NCR} = \text{angle NCA}$.

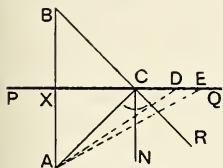


FIG. 22/13

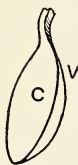


FIG. 22/14

Draw AX perpendicular to the mirror, and produce AX to RC to meet at B. You should have no difficulty in proving that the triangles ACX and BCX are congruent, that $\text{XB} = \text{XA}$ —i.e. a ray of light AC, after reflection in the mirror, travels in a direction that (produced backwards) passes through the fixed point B, where B is so placed that $\text{XB} = \text{XA}$ and that AB is perpendicular to PQ.

But what has been proved to be true with regard to AC is true of any other ray, such as AD or AE.

Therefore *all* rays coming from A, after reflection, will travel as though they came from B—i.e. the image of A is at B so placed that AB is at right angles to the mirror, that B is as far behind the mirror as A is in front of it.

So far the reflections we have considered have been from plane surfaces only. We must now consider what happens when the surface is curved.

In our earlier book we saw that a curved surface may be concave or convex. We can easily remember which is concave, because such a surface presents a *cave* or hollow to the observer. The inside of a spoon (C in Fig. 22/14) would be an example.

If a curved surface bulges towards the observer (V in the same figure) it is said to be convex.

A concave or convex *spherical* surface is one which forms part of a sphere, and it is with such surfaces (forming mirrors) that we shall be chiefly concerned in this chapter.

We must recall one or two other terms. The *centre of curvature* of a mirror is the centre of the sphere of which the mirror forms a part (C in Fig. 22/15). The middle point

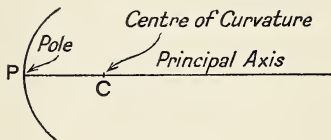


FIG. 22/15

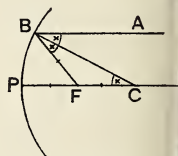


FIG. 22/16

the surface is known as the *pole* of the mirror, and the line joining the pole of the mirror to the centre of curvature, or its produced, is known as the *principal axis*.

Principal focus. Suppose a ray of light AB , travelling parallel to the axis, strikes a concave mirror at B (Fig. 22/16). Although the mirror is curved, we may take it that an *exceedingly small* portion at B is flat, its direction being at right angles to the radius BC . Thus by the ordinary law of reflection the ray AB will be reflected along BF , where $\angle CBF = \angle CBA$. Suppose this reflected ray cuts the principal axis in F . Then, by parallels, $\angle FCB = \angle CBA$.

But as already seen, $\angle CBF = \angle CBA$

$$\therefore \angle CBF = \angle FCB$$

$$\therefore FB = FC.$$

Now if AB is very close to the axis, B will be very close to P , and FB will be very nearly equal to FP .

But

$$FB = FC$$

$$\therefore FP = FC.$$

i.e. F is the middle point of CP .

Thus the ray AB after reflection passes through the middle point of CP . The same thing can be proved of *all* rays which, before striking the mirror, were parallel to the principal axis.

so all such rays are brought to a *focus* at F. This point is known as the **principal focus** of the mirror.

Notice again, however, that the rays we have been speaking of—rays parallel to the principal axis—will pass through *only if they were originally close to the principal axis*.

Caustic Curves. A pretty and interesting experiment can be carried out by making use of rays which are *not* all close to the principal axis. The mirror ACB may conveniently consist of a rectangular sheet of bright metal bent into a semi-circle. It stands upon a sheet of paper. Facing it at a little distance is a ray-box with a graining glass in front of the lens. The lamp is moved forwards or backwards until the light lines on the paper are parallel. The parallel 'rays', EG, HK, etc., are reflected by the semi-circle of metal, and give rise to a 'caustic curve' on the sheet of paper.

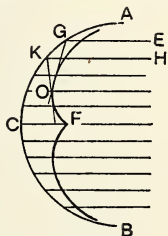


FIG. 22/17

Let us think of the point O at which two consecutive 'rays,' EG and HK, after being reflected from the mirror, cross one another. Because rays are crossing at O there is a concentration of light there—it is a sort of focus. Other pairs of consecutive rays form other foci, and it can be found by careful drawing that all these foci lie along a peculiar curve—the *caustic curve* of which we have just seen. All the reflected rays are tangents to it. It consists of two parts with a specially bright 'cusp' or prominence. This cusp is the focus of those rays which, before reaching the mirror, were travelling *close* to the principal axis. It is the 'principal focus' of which we spoke in the last chapter.

You must often have seen a caustic curve on the surface of a cup of tea, the circle of the cup in this case serving as a mirror.

Focal length. Going back to Fig. 22/16, we may notice that the distance from C to P is known as the *radius of curvature* of the mirror, while the distance FP is its *focal length*. From what has been said it will be clear that the focal length is equal to half the radius of curvature.

Various methods are in use for finding the focal length of a concave mirror. One of them depends on the fact that rays of light reaching the mirror from a very *distant* object are practically parallel. For instance, if in Fig. 22/18 the rays aA , bB . . . eE have come from a point in the sun 90 million miles away, it is clear that Aa , Bb . . . Ee would have to be produced 90 million miles before meeting—to all intents and purposes they are parallel. In most practical work with mirrors and lenses, rays may be regarded as parallel if they originate from points even 10 or 20 ft. away.

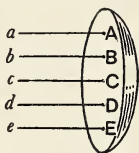


FIG. 22/18

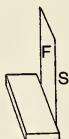


FIG. 22/19



Suppose then (Fig. 22/19), we place our mirror facing the sun, and put a screen S in front of it. We move S nearer to the mirror or farther from it until an image of the sun is sharply focused on it, at F . This image is evidently the point (or rather small area) through which rays of light parallel to the principal axis are being reflected—i.e. it is the principal focus. By measuring the distance from F to P (the pole of the mirror) we have the focal length.

The screen may consist of a post-card fastened by drawing-pins to a block of wood. A piece of three-ply painted white is better because it is more rigid.

To mark the position of P , pass one end of a knitting-needle through a cork which is gripped by the clamp of a retort stand. Make the other end of the needle touch the pole of the mirror. The mirror may then be removed, and it is easy to measure from the screen to the end of the needle.



FIG. 22/20

Instead of using the sun as our distant object, we may use the window-frames of the laboratory.

We can also use the ray-box to give us our beam of parallel rays. The method (using a graining comb) has already been

described on p. 233, when we were considering the subject of caustic curves.

The most suitable source of light is a frosted electric lamp. When it is in the right position for giving parallel rays, remove the comb and slide a sheet of glass into the groove just in front of the lens. The glass is painted white, except for a small triangle near the centre. This triangle serves as a brightly illuminated object giving off parallel rays of light.

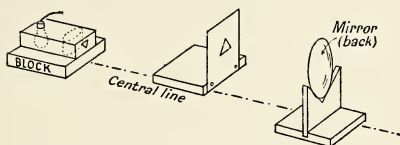


FIG. 22/21

Draw a chalk line on the bench so that the ray-box lies centrally over it as in the figure. Place the mirror in its stand and slightly to one side of this chalk line, with the postcard (mounted on its wooden block) slightly to the other side. Now move the card, with its block, until a sharply focused image of the triangle (inverted) is obtained. The distance from the mirror to the image gives the focal length.

Finding the radius of curvature. The principle used in finding the radius of curvature is a very simple one. Suppose a tiny source of light is placed at C, the centre of curvature. All the rays of light from C which reach the mirror strike it *normally* i.e. at right angles to that portion of the surface which they strike. They are therefore reflected along the path by which they came, and so the *image* of C coincides with C itself.

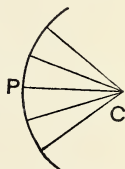


FIG. 22/22

This suggests a means of finding the position of C. We have then only to measure the distance from C to the pole of the mirror.

To carry out the experiment, put a sheet of ground glass (painted white except for a small triangle), in front of the ray-box, from which the lens has been removed. The box will be placed centrally on the chalk line, the mirror with

its stand being also arranged centrally. Now move the mirror to and fro along the line until a sharp (inverted) image is obtained, point to point with the original triangle. The

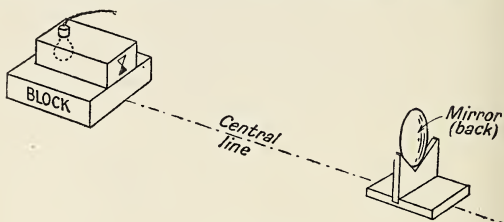


FIG. 22/23

distance from this point to the pole of the mirror gives the required radius of curvature. If we call this r , and the focal length (found in the previous experiment) is f , we ought to find, of course, that $r = 2f$.

Real and virtual images. The images so far obtained from our concave mirrors have been *real*. They were formed by the coming together of actual rays of light, and could therefore be received on a screen. We shall now meet with another kind of image.

Knowing the focal length of the mirror, put it at a distance (from the front of the ray-box) slightly less than this focal length. Now try to obtain an image on a post-card. You

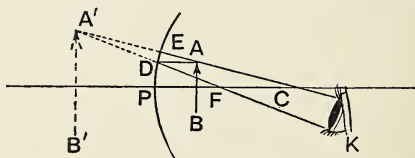


FIG. 22/24

soon find this is impossible. If, however, you stand well back from the mirror and look into it, you can see a very clear erect image. Let us try to work out what has happened.

In Fig. 22/24 P represents the pole of the mirror, C is

centre of curvature. F, bisecting PC, is the principal focus. B, on the 'mirror' side of F, stands for the illuminated triangle. (N.B. It is usual, and convenient, to represent the illuminated object by an arrow. It must not be thought that the arrow head stands for the triangle.)

From A draw a ray AD parallel to the axis CP, cutting the mirror at D. This ray will be reflected through F (p. 232). Next draw AE in the direction of CA produced—i.e. in the direction of the radius. This ray will be reflected along its own path. Clearly the reflected rays EA and DF never meet, and so cannot give rise to a real image.

But when these rays enter the eye at K, they seem to come from A' (where they meet if produced backwards). Thus A is seen at A', and similarly B is seen at B' (for clearly the construction for B' has been omitted). Hence the image of AB is seen at A'B'. Evidently rays of light are not really coming from points on A'B'; they only *appear* to be doing. Hence A'B' is known as a *virtual image*. An essential difference between the two kinds of image is that *real image can be received on a screen, while a virtual image cannot*.

Various positions of object. We have seen that a change in the position of the object causes a change in that of the image, and may also cause other changes such as erect to inverted, real to virtual, etc. Let us summarise the results so far obtained.

Position of object	Position of image	Nature of image
Very far from mirror.	At F.	Real, inverted, diminished.
At C (centre of curvature).	At C.	Real, inverted, same size.
Between F and mirror.	Behind mirror.	Virtual, erect, enlarged.

One other case remains to be explored, *viz.* when the object is between F and C. It will be found in this case that the image is on the other side of C. It is real, enlarged and inverted. It is worth noting that erect images are always virtual, and vice versa.

Relation between u , v and f . It seems reasonable to suppose that if an object is placed at a given distance in front of mirror of given radius of curvature, it should be possible to *calculate* the position of the image. Let us consider the problem. Suppose O is a point on the principal axis PCO of a concave mirror APB , whose centre of curvature is C . A ray of light from O strikes the mirror at E . Join CE . Then the reflected ray EO' cuts the principal axis in O' where $\angle O'EC = \angle OEC$. Again, a ray of light OCP , because it is travelling along a radius, will be reflected along its own path. Thus the reflected rays EO' and PO' meet at O' , which is therefore the image of O .

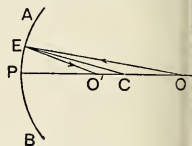


FIG. 22/25

Now because $\angle O'EC = \angle OEC$

$$\frac{CO'}{CO} = \frac{EO'}{EO} \quad \dots \dots \dots (i)$$

Again, if EP is small, then approximately

$$EO' = PO' \text{ and } EO = PO$$

\therefore instead of (i) we may write

$$\frac{CO'}{CO} = \frac{PO'}{PO} \quad \dots \dots \dots (ii)$$

Suppose OP , the distance of the object from the mirror is equal to u , and $O'P$, the distance of the image, is v . Let CP the radius of curvature $= r$.

$$\begin{aligned} \text{Then} \quad CO' &= CP - O'P = r - v \\ \text{and} \quad CO &= OP - CP = u - r \end{aligned}$$

$$\therefore \text{ equation (ii) becomes } \frac{r - v}{u - r} = \frac{v}{u}.$$

Multiplying across, $ur - uv = uv - vr$.

$$\therefore ur + vr = 2uv.$$

Divide throughout by uvr ,

$$\text{then} \quad \frac{1}{v} + \frac{1}{u} = \frac{2}{r} \quad \dots \dots \dots (iii)$$

We have seen (p. 233) that if f is the focal length, then $r = 2f$.

\therefore equation (iii) becomes

$$\frac{1}{v} + \frac{1}{u} = \frac{2}{2f}$$

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f} \quad . \quad . \quad . \quad . \quad . \quad . \quad (\text{iv})$$

Equation (iv) must be carefully remembered. It evidently enables us to calculate the position of the image provided we know that of the object, and either the focal length or the radius of curvature of the mirror.

Example 1. An object is placed 40 cm. from a mirror whose radius of curvature is 16 cm. Find the position of the image.

here $f = \frac{1}{2}r = 8$ cm., $u = 40$ cm., $v = ?$

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

$$\therefore \frac{1}{p} + \frac{1}{40} = \frac{1}{8}$$

$$\therefore \frac{1}{v} = \frac{1}{8} - \frac{1}{40} = \frac{1}{10}$$

$$\therefore v = 10.$$

e. the image will be **10 cm. from the mirror.**

Another method of finding f . The equation just considered suggests another method of finding f . All we have to do is put an object in front of a mirror and in some way find the position of its image. We then measure u (distance of object) and v (distance of image) and calculate f .

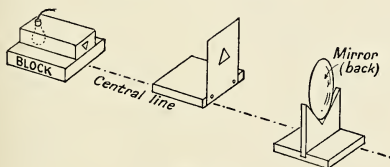


FIG. 22/26

on p. 235 we use the ray-box and a sheet of ground painted white except for a small pattern (such as a triangle). This pattern will, of course, be brightly illuminated. A mirror is placed slightly to one side of the chalk line and a screen to the other, the latter being moved towards the

mirror or away from it until a sharp (inverted) image obtained. A number of results may be obtained and tabulated thus:

u	v	$\frac{1}{u}$	$\frac{1}{v}$	$\frac{1}{v} + \frac{1}{u}$	f

With careful work it will be found that the values of f agree closely, and we should take the mean. If f has been found by one of the earlier methods (e.g. p. 234), we can evidently use the experiment just described to verify the formula

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}.$$

Magnification. In such work as the foregoing we soon notice that, generally speaking, the image and object are of the same size. The ratio

$\frac{\text{length of image}}{\text{length of object}}$ is known as the *magnification*.

There is a very simple relation between the quantities v and the magnification, which can be proved geometrically as follows. Suppose AB is the object, DPH a concave mirror, PFCB the principal axis. AD is a ray of light parallel to the principal axis, and we have already seen that this will be reflected through the principal focus F (where F bisects PC). ACH, a ray of light passing through the centre of curvature C, will be reflected along its own path. The two reflected rays meet in A', which is therefore the image of A. From A' draw A'B' perpendicular to PB.

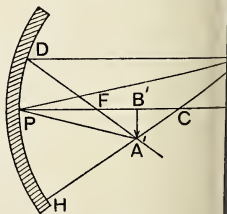


FIG. 22/27

Then A'B' will be the image of AB.

Join AP.

Then because A' is the image of A, a ray of light AP will be reflected along PA'.

These rays make equal angles with the normal

$$\angle A'PB' = \angle APB.$$

∴ the triangles A'PB' and APB are similar.

$$\frac{A'B'}{AB} = \frac{PB'}{PB}$$

$$\frac{\text{length of image}}{\text{length of object}} = \frac{\text{distance of image}}{\text{distance of object}},$$

we put I and O for length of image and object respectively,

$$\text{magnification } \frac{I}{O} = \frac{v}{u}.$$

Practical applications. The things we have learnt about plane mirrors have a number of useful applications in everyday life. Let us go back to the fact that rays of light travelling parallel to the principal axis are reflected through the principal focus. Now, by putting a source of light at F

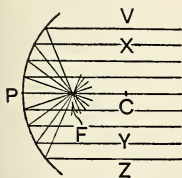


FIG. 22/28

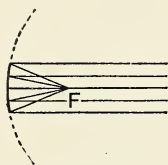


FIG. 22/29

quite easy to reverse the state of affairs. Light will then come from F, strike the mirror, and be reflected parallel to the principal axis (Fig. 22/28). Thus by putting a concave mirror in a bicycle lamp or in the head-light of a motor car or in a searchlight, and placing a small source of light at the principal focus, we can cause a parallel beam to be sent out. Such a beam will be effective for very long distances

because, unlike the ordinary 'divergent' beam, it is weakened by spreading out as it travels farther from the mirror.

We saw, however, that rays parallel to the principal axis are reflected through F only if they are *close* to the axis. Reversing the direction, it follows that only that portion of the mirror which is close to the axis will serve to give us a parallel beam. In short, it looks as though we can only use the central portion of the concave reflector, and all the other part would have to be cut away or blacked out (Fig. 22/29). In that case most of the light from F would not be reflected at all, and our parallel beam would be greatly weakened.

The difficulty is overcome by using a reflector of the shape shown in Fig. 22/30 (a 'parabolic' reflector). It can be shown that with one of this shape, *all* rays from the principal focus are reflected parallel to the principal axis.

A shaving-mirror is often made to be slightly concave (i.e. to have a large radius of curvature). A man standing between F and the mirror will then see an enlarged, erect, virtual image of his face (cf. Fig. 22/24).

A dentist also uses a concave mirror to get a view of the back part of a tooth. In this case the distance from mirror to tooth is very small, and so the focal length of the mirror need only be small (again, cf. Fig. 22/24).

Convex Mirrors. We may now try to perform experiments with a convex mirror similar to those done with concave mirrors. We soon find, however, that we cannot obtain a real image on the screen, because the images are in every case *virtual* (they are also erect, and diminished).

We will try, by drawing, to find the position of the image formed by a convex mirror, but first of all let us see

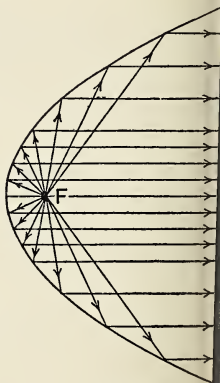


FIG. 22/30

happens to a ray of light travelling parallel to the principal axis. As before, C is the centre of curvature and P the pole of the mirror (Fig. 22/31). CP (produced to R) is the axis. AB is a ray of light travelling parallel to RP and striking the mirror at B. Join CB and produce it to Q.

As before, the radius CB may be considered to be at right angles to the tiny bit of mirror immediately surrounding B. Thus CBQ is a normal to this portion of the mirror and the ray of light will be reflected along BG, where $\angle QBG = \angle QBA$.

Produce GB to meet CP at F.

By parallels, $\angle QBA = \angle BCF$.

Also $\angle QBG = \angle FBC$
(vertically opposite)

But $\angle QBA = \angle QBG$

$\therefore \angle BCF = \angle FBC$

$\therefore FB = FC$.

If B is close to P, then (approximately) $FB = FP$.

$\therefore FC = FP$.

Thus the reflected ray BG, when produced backwards, cuts the axis at a point F such that F bisects PC. This result is, of course, similar to what we obtained with the concave mirror.

We should now have no difficulty in finding, by drawing,

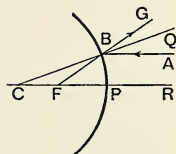


FIG. 22/31

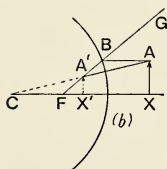
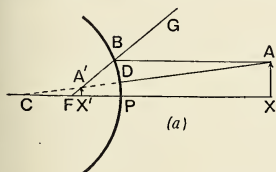


FIG. 22/32

position of an object in a convex mirror. In Fig. 22/32a, for instance, AX is the object, and we begin by marking F, halfway between C and P. The ray AB striking the mirror is reflected along BG, obtained by joining FB and producing to G. The ray AD, in line with the radius, is reflected
R

along its own path. DA and BG appear to come from X' , which is therefore the virtual image of A. X' , the image X, lies on the axis, vertically below A. In Fig. 22/32b the work is repeated, but with AX closer to the mirror.

Images formed by convex mirrors are *always* virtual, erect and diminished.

Sign Convention. In our proof of the formula $\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$ we used a concave mirror, and the object was at such distance that a real image was obtained. It can be shown however, that the formula applies in *all* cases—whether we are dealing with a concave or convex mirror, and whether the image is real or virtual—provided certain simple rules are observed. Various sets of rules or *conventions* are in use, but in this book we shall employ what is known as the *real-is-positive* convention. According to this,

1. We measure all our distances from the pole of the mirror as starting point.
2. We regard the distance as positive if it is measured to a real object, image or focus, but negative if measured to a virtual image or focus.

By studying the worked examples which follow, the reader will soon learn how these simple rules are applied.

Example 2. By a ray diagram, and also by calculation, find the position and size of the image formed when an object 2 inches high is placed 3 inches in front of a convex mirror of focal length $1\frac{1}{2}$ inches.

The ray diagram is shown in Fig. 22/33, vertical measurements (e.g. length of object) being drawn half-scale as compared with horizontal. Where (as in this case) the object is not small in comparison with the other measurements, it is best to place it symmetrically across the axis, as shown.

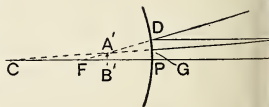


FIG. 22/33

The usual two rays from A are shown (i) AD parallel to the principal axis (reflected ray travels as though it had passed through F) and (ii) AG travelling towards centre of curvature (reflected along its own path). Thus A' is found. It is sufficient to place B' symmetrically below the axis.

By measurement A'B' is found to be 1 in. behind the mirror and to be $\frac{2}{3}$ in. long.

By calculation. Mirror is convex, \therefore focus is *virtual*, making negative in accordance with the real-is-positive convention

\therefore in formula
$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

we have $v = ?$ $u = 3$, $f = -1\frac{1}{2}$

\therefore
$$\frac{1}{v} + \frac{1}{3} = \frac{1}{-1\frac{1}{2}}$$

This gives $v = -1$, i.e. image is 1 inch from the pole of the mirror and *virtual*; so it is behind the mirror.

Magnification
$$= \frac{v}{u} = \frac{1}{3}$$

\therefore length of image $= \frac{1}{3}$ of 2 in. $= \frac{2}{3}$ in.

Thus image is $\frac{2}{3}$ in. long and 1 in. behind the mirror.

Note that in working out the magnification we ignore the negative sign of v . The sign only enables us to decide that image is virtual, and therefore behind the mirror.

Convex mirror as driving-mirror. The driver of a motor is required by law to fit his car with a mirror so that he may have a view of the road behind him. In practice he always uses a *convex* mirror, because it gives him a very wide field of view. In Fig. 22/34 suppose the driver's eye is at E,

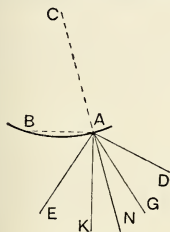


FIG. 22/34

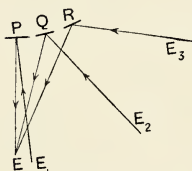


FIG. 22/35

that A is a point near the outer edge of his mirror. C is centre of curvature. Draw the normal CAN, and make $AD = \angle NAE$.

Thus the outer edge of his mirror is reflecting rays of light from points as far out as D. His field of view is represented approximately by the large angle EAD.

Now suppose he had used a plane mirror, represented by the dotted line AB. If we draw the normal AK, and make $\angle KAG = \angle KAE$, we see that his field of view is represented by the much smaller angle EAG.

We can think of the convex driving-mirror in another way. We may imagine it made up of a number of very small plane mirrors, P, Q, R. . . .

In Fig. 22/35 the incident ray E_1P corresponding to the reflected ray PE has been drawn, and similarly the incident rays E_2Q , E_3R . Thus the first mirror enables him to see objects almost immediately behind him, while the others enable him to see objects more and more to his off side.

A driving-mirror is mounted on an adjustable metal arm, and if the motorist requires a wider field of view on the off side, he pushes this arm slightly away from him.

Any car driver, by the way, will tell you that when overtaking car has nearly drawn level, its image in the driving mirror becomes much larger. This can easily be understood by a comparison of Figs. 22/32 (a) and (b) on p. 243. In (b) the object is much closer to the mirror than in (a), and you see that the image is much larger.

Similarly, when looking into a convex mirror you have doubtless been impressed with the noble proportions of your own nose. The explanation, of course, is the same as that just given. Your nose is nearer to the mirror than is the rest of your face, and so forms a relatively larger image.

We will conclude this chapter by working one or two more typical examples.

Example 3. *An electric-lamp bulb is placed with its very small filament on the principal axis of a concave mirror, and 4 cm. in front of it. The reflected light appears to diverge from a point which is 12 cm. behind the mirror. What is the radius of curvature of the mirror? J.M.B.**

The image is virtual, and 12 cm. from the mirror.

Thus in the formula $\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$ we have

$v = -12$ cm., $u = 4$ cm., and $f = ?$

$$\therefore \frac{1}{-12} + \frac{1}{4} = \frac{1}{f}$$

This gives $f = 6$ cm. and the radius of curvature $= 2f = 12$ cm.

Example 4. Find the distance from the face at which a concave mirror must be held to give an image magnified two-fold, the radius of curvature of the mirror is 4 ft. O.C.*

Here the image is obviously erect. It must therefore be behind the mirror, and virtual, i.e. v is negative.

$$\text{Now} \quad \frac{1}{O} = \frac{v}{u}, \text{ and } \frac{1}{O} = 2$$

$$\therefore \quad \frac{v}{u} = 2 \text{ (numerically)}$$

$$\text{But } v \text{ is negative, } \therefore v = -2u$$

$$\text{Also} \quad r = 4 \text{ ft.}$$

$$\therefore \quad f = 2 \text{ ft.}$$

$$\text{If required distance from face} = u \text{ ft.}$$

$$\text{Then} \quad \frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

$$\frac{1}{-2u} + \frac{1}{u} = \frac{1}{2}$$

$$\text{giving } u = 1 \text{ ft.}$$

Example 5. Taking the diameter of the sun to be 870,000 miles, the distance of the sun from the earth 93,000,000 miles, find the diameter of the image of the sun produced by a concave mirror of radius of curvature 100 ft. O.C.*

Since the object is very distant, the image will be formed at principal focus, which is 50 ft. (half the radius of curvature) in front of the mirror.

$$\therefore \quad v = 50 \text{ ft.} = \frac{50}{5280} \text{ miles.}$$

$$\text{Let required diam.} = I \text{ miles.}$$

$$\text{Then magnification} = \frac{I}{O} = \frac{v}{u}$$

$$\therefore \quad \frac{I}{870,000} = \frac{\frac{50}{5280}}{93,000,000} = \frac{50}{5280 \times 93,000,000}$$

$$I = \frac{50 \times 870,000}{5280 \times 93,000,000} \text{ miles}$$

$$= \frac{50 \times 870,000}{5280 \times 93,000,000} \times 5280 \text{ ft.}$$

$$= 0.47 \text{ ft.} = 5.6 \text{ in.}$$

Questions

A

1. How are shadows formed? Draw a diagram to show the kind of shadow you would expect to find on a screen when a large spherical source of light casts a shadow of a small ball upon it. What fact do you make use of in your drawing, and how would you verify it experimentally? *Dur.*

2. State the laws of reflection of light, and describe an experiment by which you could prove one of them.

3. Make a labelled diagram of a concave mirror showing the *centre of curvature*, the *principal focus* and the *principal axis*. Also show the *pole* of the mirror.

4. Use diagrams to illustrate the formation of *real*, *virtual*, *erect* and *inverted* images by means of concave and convex mirrors. *O.C.*

5. What is meant by the terms *radius of curvature*, *focal length*, *virtual image*?

A convex mirror has a radius of curvature of 20 cm. Draw a diagram to scale to show the paths of rays of light from an object which gives rise to an image of half the size. What is the distance of the object from the mirror? A graphical solution of the problem would be preferred. *Dur.*

6. A luminous object is placed 6 in. from (a) a plane mirror, (b) a concave mirror of focal length 9 in. In each case draw a ray diagram to illustrate the formation of the image, and give details of the nature, position, and magnification of the image. *C.W.B.**

7. By means of a ray diagram, explain the working of a motor car driving-mirror.

B

1. Describe the formation of an image by a pin-hole camera. Explain what happens to the image when (a) the hole is gradually enlarged, (b) the screen is moved away from the pin-hole, (c) the shape of the hole is altered.

What fundamental principle of optics is illustrated by these experiments? *Camb.*

2. If the length of a pin-hole camera from hole to plate is 2 in., and a 6-ft. man stood 9 ft. from the hole, what would be the size of his picture? *J.M.B.**

3. Explain briefly, using diagrams, how an eclipse of the sun takes place, distinguishing between total and partial eclipses. What law of light do you assume in your explanation? *Camb.*

4. Describe an experiment by which you could show that if a small object is placed in front of a plane mirror, its image is as far behind the mirror as the object itself is in front of it.

6. Assuming the laws of reflection of light, give a geometrical proof of the statement in question 4.
7. Describe an accurate experiment to find the focal length of a concave mirror.
8. An object, 3 cm. high, is placed 15 cm. from the pole of a concave mirror of 10 cm. radius of curvature. Find the position of the image by a graphical construction, which should be done full scale. Explain each step in the construction. State also size and nature of the image. *Oxf.*
9. Find the position and size of the image of an object 4 cm. high placed 100 cm. in front of a concave mirror of focal length 25 cm. *Camb.**
10. An object is 10 in. from the centre of curvature and 20 in. from the principal focus of a concave mirror. What is the position and magnification of the image? What changes in the image occur when the object is moved a distance of 25 in. towards the mirror? *Dur.*
11. By means of a ray diagram, explain the use of a spherical concave mirror to give an erect image, magnified twice.
12. A man using a concave shaving mirror sees an erect image of his face enlarged one and a half times when his face is 10 in. from the mirror. Draw a diagram to scale to illustrate the formation of the image, and from it deduce the distance of the object from the mirror and the radius of curvature of the mirror. *Dur.**
13. Describe the optical arrangement in a searchlight giving a parallel beam, and explain the adjustment necessary to give a converging beam.
14. Every day, searchlights should never be left in such a way that they might point at the sun; explain this fact. *J.M.B.**
15. How would you verify the formula relating the distance of an object and its image from the pole of the mirror?
16. The following points, plotted on squared paper, will give a parabola with its focus F at $x = 0.5$.

x	0	0.1	0.2	0.5	1.0	1.5	2.0
y	0	0.45	0.63	1.0	1.41	1.73	2.0

Draw the parabola. Regarding F as a point source of light, let any ray to meet the curve at A_1 , and from A_1 draw a line A_1P parallel to the axis. This should be the reflected ray corresponding to the incident ray FA_1 . To test this, draw a tangent to the curve at A_1 (by eye, using your ruler), and then draw the normal at A_1 . Measure the angles of incidence and reflection.

Repeat with several other points, A_2, A_3 , etc.

CHAPTER 23

PHOTOMETRY

IN the preceding chapter there was a reference to the measurement of the velocity of light. In the present one we shall consider some much more 'every-day' measurements. If a man were buying a source of light (e.g. an electric light bulb) he would not bother about the speed at which light from the bulb travels into space, but he would certainly want to know how much light it throws out—i.e. its *illuminating power*.

Again, a man in a reading-room may complain to the attendant that the lighting is poor, and the attendant may reply that the lights in use are actually very powerful. Both reader and attendant may be right. The powerful lights may be at too great a distance from the periodicals, etc. We see that what is wrong is not the illuminating power, but the *intensity of illumination*. We must evidently distinguish clearly between the two. Illuminating power is a quantity that depends only on the source of light, while intensity of illumination depends not only on the strength of the source of light, but also on its position. The distinction will become clearer as we proceed.

The Standard Candle. Brightness. To measure illuminating power we evidently need some unit, and the one chosen is the *standard candle*. This was defined by an Act of Parliament passed in 1860 as *a candle made of sperm wax, weighing one-sixth of a pound, and burning away at the rate of 120 grains per hour*.

Of course the important point about a unit is that it should be *constant*, and it was thought that by defining the material and the rate of burning, etc., a really constant illuminating power would be secured. Actually the standard candle is not a very satisfactory unit, for its illuminating power varies slightly according to the shape of the wick, the pressure and humidity of the air and other circumstances. Vernon Harcourt found that much more constant results could be obtained by burning pentane vapour under certain specified conditions (pentane is a light oil obtained from paraffin). The 'Harcourt

acetylene lamp' has an illuminating power equal to that of standard candles, and is now always used as the official standard.

The candle-power of a few well-known sources of light may be mentioned. A 50-watt lamp has usually a candle-power of about 40; a good incandescent gas mantle about 96. Electric torches naturally vary a great deal, but a very common type (3.5 volts, 0.3 amps) is from 1 to 2 c.p.

We must not confuse candle-power with *brightness*. Suppose an incandescent gas mantle and an acetylene flame are of equal candle-power (as might quite well be the case) but the mantle has a surface area of 10 sq. in. and the acetylene flame only 1 sq. in. The acetylene flame would be much more dazzling to look at than the mantle. That is because the amount of light given off *per square inch of surface* is ten times as great.

It is worth noticing that it is this factor—light given off *per unit area*—which determines brightness. That is why the modern electric-lamp filament is so often enclosed in translucent instead of transparent glass. Owing to the smallness of the filament, the light is given off from a very small area, and the filament would be too dazzling for comfort. Earlier lamps were of clear glass. They did not (necessarily) give off less light, but the light was given off from a stouter filament, which was therefore less brilliant than the modern one, with its smaller surface area.

Intensity of Illumination. Let us now look more closely at the question of intensity of illumination, which, as we have seen, depends both on the candle-power of the light (or lights) concerned and on its distance from the surface we are considering. A fairly obvious unit is the **foot-candle**, which we define as *the intensity of illumination of a surface at one foot from a standard candle, and in such a direction that the light strikes it at right angles*. The last half of the definition is often overlooked, but it is important, because the intensity with which a surface is illuminated depends very much on the direction at which the light strikes it. This point will be taken up again in Example 3 (p. 260).

For comfortable reading we need an intensity of illumination of at least 1 ft.-candle, and a woman sewing with dark material

would need 6 or 7. A watchmaker would need still more—10 or 12. The lighting of a classroom would be considered poor if the intensity of illumination were less than 3 ft candles. We shall see presently how these intensities can be measured, but we cannot attack this problem until we have obtained an answer to a question which we may express thus: *How much* does the intensity of illumination of a surface fall off as we increase its distance from the source of light?

For instance, light from a standard candle is falling upon a surface 1 ft. away. The intensity of illumination is obviously 1 ft.-candle. What would this intensity be if we increased the distance to 2 ft. or to 3 ft.? (It is understood that the light falls 'normally' upon the surface in all cases.)

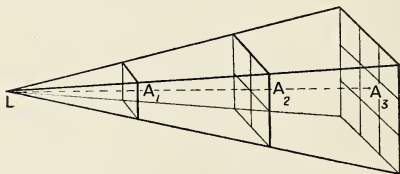


FIG. 23/1

Suppose a small source of light L is throwing out a beam of square cross-section. (If L is contained in a glass enclosure this is easily secured in practice by pasting black paper on the inside of the container, and then cutting out a small square.) Suppose the square A_2 is twice as far from L as is the square A_1 . Because the distance is twice as great, it follows from similar triangles that a *side* of the square A_2 is twice as long as a *side* of the square A_1 . Therefore the *area* of A_2 is four times the *area* of A_1 . As the same amount of light reaches A_2 as A_1 , but has to cover four times the area, it follows that the intensity of illumination of A_2 is only a quarter that of A_1 .

Similarly, if the square A_3 is three times as far away as A_1 , the intensity of illumination of A_3 will be only one-ninth that of A_1 and so on.

The conclusion thus reached is known as the **Law of Inverse Squares**, and may be stated as follows: *The intensity of illumination is inversely proportional to the square of the distance from the source.*

illumination of a surface varies inversely as the square of distance from the source of light.

Thus, while the intensity of illumination due to a standard candle at a distance of 1 ft. is 1 ft.-candle (by definition), at 2 ft. the intensity is only $\frac{1}{4}$ ft.-candle, at 3 ft. $\frac{1}{9}$ ft.-candle and so on. If we use two candles, the intensities (in foot-candles) would be respectively 2, $\frac{2}{4}$ ($= \frac{1}{2}$) and $\frac{2}{9}$. In general, if c = illuminating power measured in candles, i = intensity of illumination (in foot-candles) and d = distance (in feet), then $i = \frac{c}{d^2}$. Thus we have a simple numerical relationship connecting illuminating power, distance and intensity of illumination.

With the law of Inverse Squares in mind, we might think that in an ordinary room the intensity of illumination at 12 ft. from the electric-light bulb would be only one-ninth of the intensity at 4 ft. (distances are as 3 : 1, therefore intensities are as 1 : 3²). This is by no means the case, because the walls of the room reflect the light. Thus every part in the room receives light, not only from the bulb, but from the walls, and the illumination may be fairly even throughout.

By this time the reader will have acquired fairly accurate ideas of the terms Illuminating Power and Intensity of Illumination, and it will be well to express them in formal definitions.

The **illuminating power** of a source of light means *the rate at which energy in the form of light is being emitted by that source*. In practice, it is measured in standard candles.

The **intensity of illumination at a point**, means *the rate at which energy in the form of light is being received per unit area immediately surrounding the point, such area being supposed to be at right angles to the direction of the light*.

It is measured in foot-candles. Another unit is the metre-candle or *lux*.

Comparison of Illuminating Powers. We are now in a position to compare the illuminating powers of two sources of light, say a small electric bulb and a gas flame.

Place a sheet of drawing-paper to the bottom of a drawer, thus forming a screen (Fig. 23/2). Place a retort stand so that the

vertical rod is 2 or 3 in. in front of the screen, and place the gas flame G (a fish-tail burner) so as to throw a shadow S_g . Now place the electric bulb B so that the shadow it makes, S_b , is almost touching S_g .

Clearly the shadow S_b is illuminated by the gas-flame, while S_g is illuminated by the bulb. Move the gas-flame (or bulb) towards or away from the screen until the shadows are of equal depth, and then measure the distance of each source of light from the shadow it is illuminating. Suppose this distance is 24.0 in. for the gas-flame (candle-power c_g) and 36.0 in. for the bulb (candle-power c_b). Then the intensities of illumination may be expressed

as $\frac{c_g}{24^2}$ and $\frac{c_b}{36^2}$ respectively. But we have arranged that the

intensities shall be equal. Hence we have $\frac{c_g}{24^2} = \frac{c_b}{36^2}$, from which we easily work out that $c_g : c_b = 1 : 2\frac{1}{4}$.

A similar experiment is sometimes carried out to illustrate the law of Inverse Squares. Five similar candles are taken and four of them, placed close together, are used as a single source of light, to be compared with the light from the remaining candle. When the shadows are of equal intensity, it is found that the '4-candle' light is at *twice* the distance of the '1-candle' light. Calling these distances $2d$ and d , we have

Intensity of illumination from 4 candles at distance $2d$
 = Intensity of illumination from 1 candle at distance d .
 \therefore Intensity of illumination from 1 candle at distance $2d$
 = 1/4th intensity of illumination from 1 candle at distance d .

Thus as the distance is doubled the intensity of illumination is reduced to one-quarter, as required by the law of Inverse Squares. Other results may be obtained by varying the number of candles. Actually the experiment is not a very

¹ 24 in. = 2 ft., so that the intensity in foot-candles would be $\frac{c_g}{2^2}$. $\frac{c_g}{2^2}$ really gives the intensity in inch-candles, but there is no objection so long as we do the same for both sources of light.

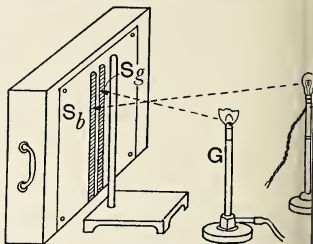


FIG. 23/2

isfactory one, partly because when candles are burning together the wax melts at an abnormal rate, and each candle gives more than its ordinary amount of light.

An instrument used for measuring candle-power (or comparing candle-powers) is known as a *photometer*, the process itself being known as photometry (cf. calorimeter and calorimetry). The very simple instrument we have so far used is called *Rumford's shadow photometer*. The inventor, Rumford, the same man we have met with already in connection with the overthrow of the caloric theory of heat.

Bunsen's Grease-Spot Photometer. Another very simple effective photometer depends on the use of a grease-spot, was invented by the famous Bunsen.

Let a drop of fat or oil fall on a piece of paper, and hold the latter up to the window. The grease-spot is lighter than the rest of the paper. Hold it away from the window. It appears dark. Clearly, when we view the grease-spot from the side of greater illumination it appears darker than the rest of the paper, and when we view it from the side of lesser illumination it appears lighter. It is reasonable to suppose that *when the illumination is equal on the two sides, the grease spot will appear neither lighter nor darker than the rest of the paper*—i.e. the grease-spot as such will cease to be visible.

We can easily test this conclusion by putting our greased paper in a simple wire frame, darkening the room, and then putting lighted candles on opposite sides of the frame, but at equal distances, so that the illumination shall be the same on both sides. The grease-spot is invisible, but immediately comes into view if we disturb the balance by changing the position of one of the candles.

It is not very hard to understand the appearance of the grease-spot. It is more translucent than untreated paper—it allows light to pass through it more readily. Suppose we are viewing it from the side on which more light is falling. Because it is transmitting *more* light than the rest of the paper, it is reflecting *less*, and so it appears darker. If, however, the illumination is the same on both sides, the grease-spot, though still *reflecting* less light than the rest of the paper, makes up the deficiency by the extra light it is

now *transmitting* from the other side. The result is that observer now receives exactly as much light from the great spot as from the rest of the paper, and so he cannot distinguish one part from the other.

Determination of Candle-power. Suppose we wish to use the photometer to compare the candle-powers of two sources of light, say the electric bulb and fish-tail burner already used with the shadow photometer.

Put the bulb on one side and the fish-tail burner on the other and adjust the distances so as to make the spot disappear.

Two mirrors are so placed that both sides of the screen can be viewed at the same time—a considerable improvement.

Suppose the distance of the bulb is found to be 30 in. and that of the burner 20 in. Let c_b = candle-power of bulb and c_g = candle-power of burner. Then intensity of illumination

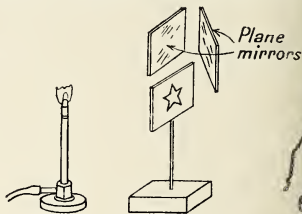


FIG. 23/3

due to bulb = $\frac{c_b}{30^2}$ and intensity of illumination due to gas-burner = $\frac{c_g}{20^2}$ (cf. foot-note p. 254).

But the intensity is the same for each side.

$$\therefore \frac{c_b}{30^2} = \frac{c_g}{20^2}$$

$$\therefore \frac{c_b}{c_g} = \frac{30^2}{20^2} = \frac{900}{400} = 2\frac{1}{4} : 1.$$

No matter what kind of photometer has been used in making the above comparison, if the candle-power of one of the sources of light is known, we can easily calculate that of the other. Thus in the case just considered, suppose the bulb is known to be of $4\frac{1}{2}$ c.p. For $\frac{c_b}{c_g} = \frac{2\frac{1}{4}}{1}$ we write $\frac{4\frac{1}{2}}{c_g} = \frac{2\frac{1}{4}}{1}$ whence $c_g = 2$ c.p.

To secure good results with the grease-spot photometer certain precautions have to be taken.

There must be no sources of light other than the two being

pared—i.e. the room must be darkened. It is desirable to have darkened walls, for otherwise, by reflecting, they are as sources of light.

Unglazed paper should be used for a screen. Owing to the power of reflection, glazed paper is very likely to be brighter in some parts than others, and this would affect the degree of visibility of the grease-spot.

Keeping one of the sources of light, and also the photometer, fixed in position, *four* readings of the distance of the other source of light should be obtained and the average of four taken as the true reading.

- (i) grease-spot invisible as viewed from one side;
- (ii) the same, but viewed from the other side;
- (iii) and (iv)—the same as (i) and (ii), but with the photometer made to face in the opposite direction.

The experiment should be repeated several times with the positions of light at various distances from the photometer, the results being worked out and the average taken as the answer.

Experiments with the shadow photometer—

- (i) The room need not be specially darkened, because stray light will affect both shadows to the same extent.
- (ii) Unglazed paper should be used.
- (iii) The question of reversing the photometer, etc., obviously does not arise.
- (iv) Repetition of the experiment, with the sources of light at various distances is, of course, desirable.

The shadow photometer is now of little more than historical interest. Bunsen's instrument is better, but many photometers have been devised which are superior to either of them. One of them—not at all hard to make—is the *Conroy Photometer*. The instrument is shown in plan (with the lid removed) in Fig. 23/4. It consists of a small box in which two openings have been made, one for observation, the other two to admit light from the two sources to be compared. The light falls on screens of *unglazed* cardboard, and must fall on them at the same angle. In practice about 58° is found to be suitable.

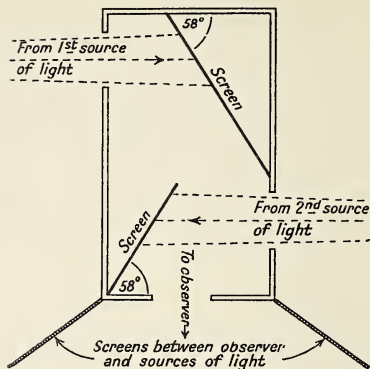
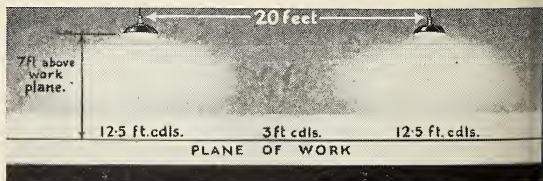
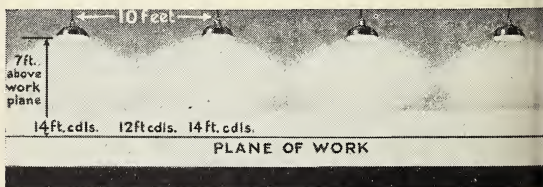


FIG. 23/4

On making a first observation it will probably be found that the edge of the front screen is easily seen against the other one as a background. This means that the two



By courtesy of Lighting Service Bureau

WHAT THE PHOTOMETER REVEALS

t equally illuminated, and the distances of the sources of light are adjusted until the edge of the front screen is practically visible. The distances from the two sources of light to the edge are then measured, and the results are worked out on the principle already explained.

It is often necessary to measure the intensity of illumination in different parts of a room, and for this a *light-meter* is employed.

The essential part of this instrument is a photoelectric cell, which depends on the principle that an electric current may be produced by the action of light on a specially prepared surface, the intensity of the current being proportional to the intensity of illumination. A needle moves over a scale graduated in foot-candles as in Fig. 23/5.

The illustration on p. 258 suggests that a light-meter may be used to compare the efficiencies of two different systems of lighting.

We may conclude this chapter with one or two calculations.

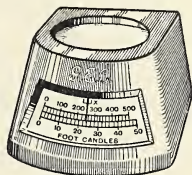


FIG. 23/5

By courtesy of General Electric Co. Ltd.

Example 1. *A man can just read a newspaper when he stands at a distance of 20 ft. from a 100-candle-power lamp. At what distance could he just read it using a single candle?*

Intensity of illumination = $\frac{100}{20^2}$ in the first case and $\frac{1}{x^2}$ in the second, where x is the required distance in feet. These intensities are equal, being in each case the minimum required for reading.

$$\begin{aligned}\therefore \quad \frac{100}{20^2} &= \frac{1}{x^2} \\ \therefore \quad 100x^2 &= 400 \\ \therefore \quad x &= 2 \text{ ft.}\end{aligned}$$

Example 2. *Two lamps of equal candle-power are placed 8 ft. apart, with a grease-spot photometer midway between them. A smoked-glass screen is now placed against one of the lamps, and it is found that it has to be moved 1 ft. nearer the photometer to make the grease-spot invisible once more. What percentage of the light is cut off by the smoked screen?*

Suppose the candle-power of each lamp = c , and that $x\%$ the light of one of them is cut off by the screen. Its candle-power may now be regarded as $\frac{100-x}{100} \times c$, and its distance from the screen has been reduced to 3 ft.

$$\begin{aligned} \therefore \text{intensity of illumination} &= \frac{(100-x)c}{100 \cdot 3^2} \\ &= \frac{(100-x)c}{900} \text{ ft.-candles.} \end{aligned}$$

The unscreened lamp (c candle-power at a distance of 4 ft.) gives an intensity of $\frac{c}{4^2} = \frac{c}{16}$ ft.-candles.

The two intensities are equal.

$$\therefore \frac{(100-x)c}{900} = \frac{c}{16}$$

$$\text{Cancelling } c, \text{ we have } \frac{100-x}{900} = \frac{1}{16}$$

$$\therefore 1600 - 16x = 900$$

$$\text{which gives } x = 43\frac{3}{4}.$$

Thus the percentage of light cut off is $43\frac{3}{4}\%$.

Questions

1. What do you understand by the inverse square law applied to optical illumination? How is it applied in the laboratory to compare the illuminating powers of lamps?

A satisfactory photographic print on gas-light paper is obtained when the printing-frame, held at a distance of 2 ft. from a small electric lamp, is exposed to the light for 5 sec. At what distance from the lamp must the frame be held in order that an exposure of 20 sec. shall produce the same result? *O.C.*

2. Describe a simple photometer and show how to use it to compare the candle-powers of two similar lamps. Why is it essential to conduct photometric experiments in a darkened room?

Calculate the intensity of illumination at a point on a screen 2 ft. away from a small 10 candle-power lamp and perpendicular to the incident light. What will the intensity of illumination on the screen be if a plane mirror is placed 6 ft. behind the lamp and parallel to the screen, assuming that the mirror reflects 100% of the light incident upon it? *O.C.*

3. Explain carefully the meaning of *intensity of illumination at a point on a surface*.

A small 20-candle-power lamp is placed 4 ft. above the centre of a circular table which it illuminates. Would the table be uniformly illuminated? Give reasons.

Find (1) the intensity of illumination at the table centre

very far above the centre the lamp should be placed to produce intensity of illumination there twice as great as it was originally? *Dur.*

. The candle-powers of two sources of light are in the ratio 4 : 1. If they are placed 200 cm. apart, where should a screen be placed so that the intensities of illumination on both sides be equal? *O.C.**

. An electric lamp is 6 ft. from a screen. A glass plate which lets off 19% of the light is interposed between the lamp and the screen. How much must the lamp be moved in order to restore intensity of illumination on the screen to its original value? *Lond.**

. A bright light gives an illumination of 2 ft.-candles 40 yd. away. Calculate the illumination at 20 yd. distance if a sheet of paper which absorbs half the light incident on it is placed in the way of the lamp. *J.M.B.**

. Define a *foot-candle*. A 40-candle-power lamp is placed 10 ft. away from another lamp L, and a screen placed between them at 5 ft. from the 40-candle-power lamp is found to be equally illuminated on both sides. If the 40-candle-power lamp is replaced by one of twice its power, the screen being kept in the same place, where must the lamp L be placed if the screen is still to be equally illuminated on both sides? *C.W.B.*

CHAPTER 24

REFRACTION AT PLANE SURFACES

NOT many pages back we were considering a very important fact about light—that it travels in straight lines. In the present chapter we shall be discussing the fact that it does not at least not always!

The contradiction, of course, is only apparent. A ray of light maintains its straight-line direction so long as it continues to travel *in the same medium*. If it enters another medium it changes its direction, the phenomenon being known as *refraction*. We have a very familiar illustration in the fact that a straight stick with its end dipping in water appears to be bent at the point where it enters the water (Fig. 24/1).

As we saw in our earlier book, this is because a ray of light passing through a certain medium (in this case water) changes its direction on entering another medium (in this case air), this change of direction being known as *refraction*.

Suppose ABC is the true position of the stick, the end C being in water. Consider a ray of light CD travelling from water to air. It is refracted at D, follows the changed direction DE, and enters the eye at E. Similarly another ray CF follows the path FG and enters the eye at G.

Produce ED and GF to meet at C'. Now, to the eye the rays DE and FG seem to come from C', and so the point C seems to be at C'. B, where the stick cuts the surface, will be seen in its true position, while points between C and B (e.g. M) will be seen in positions between C' and B (e.g. M'). Thus BC will be seen in the position BC', and the submerged part of the stick will appear 'bent.'

Other facts can be explained in a similar way.

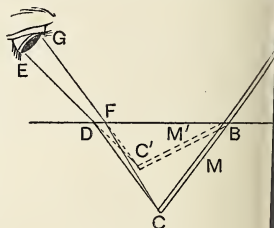


FIG. 24/1

Fig. 24/2a shows an observer's eye so placed that he cannot see the brass plug at the bottom of a wash-bowl. In Fig. 24/2b the plug has come into view, though his eye is still in the same position.

In Fig. 24/3, C is (say) a pebble at the bottom of a pond. Owing to refraction, rays of light from C appear to come from

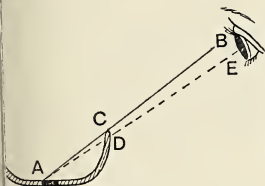


FIG. 24/2a

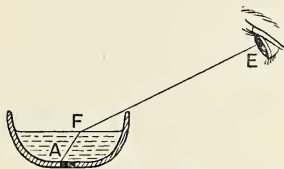


FIG. 24/2b

nearer the surface—i.e. the pond appears to be shallower than it really is.

N.B. It will be seen later that KC' is about three-quarters KC , and this suggests the best way to draw a good diagram. For marking C, put in C' at three-quarters the depth. Draw straight lines $C'DE$ and $C'FG$ (to reach the eye at E and G), meeting the surface at D and F. Then join CD , CF . (In the figure on the right, the effect is exaggerated for the sake of clearness. The figure on the left is more accurate.)

In the experiments so far discussed, we cannot actually see the rays of light changing direction as they pass from water to air, as shown in Figs. 24/1-3. It is quite easy, however, to make this change of direction visible. With the help of a cylindrical lens, the ray-box is able to give a parallel divergent beam. Into the water at the front of the tank we slide a piece of

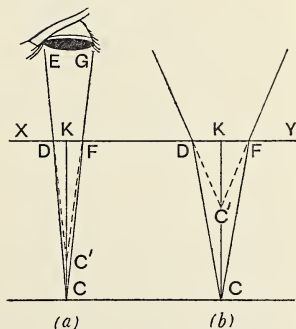


FIG. 24/3

blackened cardboard, with a hole of about $\frac{1}{4}$ -in. diameter in the middle of it. The beam strikes a mirror tilted as shown in Fig. 24/4. This throws it obliquely downwards, and strikes the upper surface of water in a trough suitably supported. The path of the beam through the water is made visible by the addition of a little eosin, and to follow its path through the air we make the latter smoky with the help of smouldering brown paper.

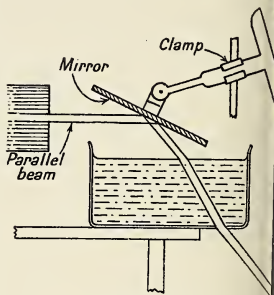


FIG. 24/4

After leaving the mirror, the beam is observed to change its direction on entering the water and again on leaving it. The directions before entering and after leaving appear to be parallel.

Measurement. Although in the preceding two or three pages we have repeatedly noted the *fact* of refraction, we have so far made no attempt at *measurement*. We can make some progress with the apparatus just mentioned. By tilt-

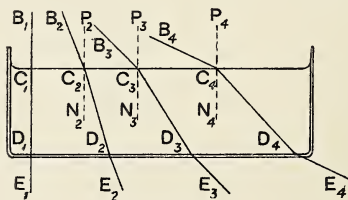


FIG. 24/5

the mirror at different angles we can vary the angle of incidence. Thus (Fig. 24/5) the incident ray B_1C_1 is travelling along the normal—i.e. the angle of incidence is 0° . In the next experiment the angle of incidence is increased to $B_2C_2P_2$, afterwards to $B_3C_3P_3$ and $B_4C_4P_4$.

As we increase the angle of incidence it is soon very obvious

As we are also increasing the angle of refraction (i.e. the angle between the refracted ray and the normal). The angle $\angle C_3D_3$, for instance, is greater than $\angle N_2C_2D_2$, and $\angle N_4C_4D_4$ than $\angle C_3D_3$. It is clear then that *the greater the angle of incidence, greater the angle of refraction*.

At once the further question arises: Is the angle of refraction *proportional* to the angle of incidence? For instance, if we double the angle of incidence, shall we find that the angle of refraction is also doubled?

Clearly we need to be able to measure our angles, and with the apparatus of Fig. 24/4 we cannot do this with any degree of accuracy. Something can be done on the lines indicated in Fig. 24/6. A slab of plate glass 'frosted' on one of its

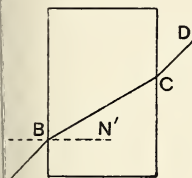


FIG. 24/6

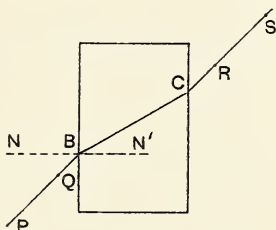


FIG. 24/7

angular faces is laid rough side downwards on a sheet of paper. AB is a very narrow beam of light from a ray box which strikes the slab at B, and we can easily trace its path through the glass. The angle of incidence $\angle ABN$ and the angle of refraction $\angle CBN'$ can be measured with some degree of accuracy. We can then vary the angle $\angle ABN$, producing corresponding changes in the other angle, $\angle CBN'$, and we can try to find if there is any definite connection between the two.

More accurate results can be obtained by the use of pins.

The slab (which need not be frosted) is laid down on a sheet of paper as before, and two pins P, Q, are placed upright on the left (Fig. 24/7). Two other pins R and S are now so placed on the other side of the block, that as we look *through* the glass (not at the top), P, Q, R and S all seem to be in a straight line.

This indicates that a ray of light which entered the glass in the direction PQ has emerged in the direction RS. With a sharp pencil, we draw the outline of the block and mark the position of the pins. Block and pins may then be removed. We join PQ and produce to meet the trace of the block at B. Similarly we obtain C by producing SR. BC now marks the path of the ray through the block.

We draw the normal NBN' and measure the angle of incidence NBP and the angle of refraction N'BC.

We now repeat the experiment, setting P, Q, at a different angle to the normal and obtaining a new value for the angle of refraction N'BC; and so on until we have got a set of values something like this:—

Angle of incidence NBP	0°*	10°	20°	30°	40°	50°	60°	70°
Angle of refraction N'BC	0°	6½°	13°	19½°	25½°	31°	35°	39°

* Notice that a ray travelling *normally* passes into the second medium without deviation.

The above results are about as nearly correct as one could expect to obtain, using an ordinary protractor. In none of them does the error exceed ¼°. Can we see any obvious connection between the angles of incidence and the angles of refraction? As the angle of incidence doubles, from 10° to 20°, the angle of refraction doubles from 6½° to 13°. As it trebles, 10° to 30°, the angle of refraction also trebles, 6½° to 19½°. It looks as if we are on the track of a very simple law. But when the angle of incidence goes up four-fold, 10° to 40°, the angle of refraction goes up from 6½° only to 25½° (though 6½° × 4 would be 26). We are just a little uneasy.

If our 'very simple law' holds good, the numbers under 50°, 60° and 70° should be 6½° × 5 = 32½°, 6½° × 6 = 39°, 6½° × 7 = 45½° respectively. Instead, they are only 31°, 35° and 39°. The law is evidently not quite so simple as it looks at first sight. We might say, perhaps, 'As the angle of incidence increases, the angle of refraction increases in the same ratio; but this statement is true only for *small* angles of incidence, up to 30° or 40°.'

This is not very satisfactory, but it seems to be the best we can do.

And indeed it was the best that *anybody* succeeded in doing.

for many centuries until 1621, when Snell, a young professor of mechanics at Leyden, hit upon the true law. Let us re-write the above results, making provision for entering the *sines* of the angles as well as the angles themselves.

Angle of incidence (i)	0°	10°	20°	30°	40°	50°	60°	70°
Angle of refraction (r)	0°	6½°	13°	19½°	25½°	31°	35°	39°
$\sin i$	0	0.174	0.342	0.500	0.643	0.766	0.866	0.940
$\sin r$	0	0.113	0.225	0.334	0.430	0.515	0.574	0.629
$\frac{\sin i}{\sin r}$		1.54	1.52	1.50	1.50	1.49	1.51	1.49

Making allowance for errors of experiment (remember our measurements were made only to the nearest half-degree), the law stands out pretty clearly. The *sine* of the angle of incidence is in a constant ratio to the *sine* of the angle of refraction. A more exact statement of the law will be made shortly.

Going back to our experiment with the narrow beam (fig. 24/6) we notice that the refracted ray lies *in* the plane of the paper—i.e. in the plane containing the incident ray AB and the normal BN. We get the same result, though a little less obviously, from the experiment with the pins. Here, instead of the narrow beam travelling visibly in the plane of the paper, we have only to remember that it travels in the *hole* made by P to the holes made by Q, R, S, all of which, of course, lie in the plane of the paper. Our laws of refraction may now be formally stated.

1. *The refracted ray lies in the plane containing the incident ray and the normal, and on the opposite side of the normal.*
2. *For a given pair of media, the sine of the angle of incidence bears a constant ratio to the sine of the angle of refraction.*

N.B. If the two media are denoted by A and B, this constant ratio is known as the *index of refraction from A to B*, and is expressed as ${}_A\mu_B$ or μ_{AB} (μ is a Greek letter pronounced 'mu'). If the first medium A is air,¹ the constant ratio is then known as the *refractive index of B*.

Strictly speaking, we should say, 'if A is a *vacuum*'; but the difference is so slight that for ordinary purposes we may ignore it.

In the table on p. 267, for instance, $\sin i \div \sin r$ (coming to about 1.50) gives the refractive index of glass.

Direction of refracted ray. If we know any two of the three quantities i (angle of incidence), r (angle of refraction) and μ (index of refraction), we can evidently use the relationship $\mu = \sin i \div \sin r$ to find the third. If for instance $i = 30^\circ$ and $r = 19\frac{1}{2}^\circ$, then $\mu = \frac{\sin 30^\circ}{\sin 19\frac{1}{2}^\circ} = \frac{0.500}{0.334} = 1.50$.

Example 1. A narrow beam of light passes from air into water the angle of incidence being 60° . Find the angle of refraction given that the index of refraction of water is 1.33.

We have $i = 60^\circ$ and $\frac{\sin i}{\sin r} = \mu$

$$\therefore \frac{\sin 60^\circ}{\sin r} = 1.33.$$

$$\therefore \sin r = \frac{\sin 60^\circ}{1.33} = \frac{0.866}{1.33} = 0.651.$$

Referring to a table of sines, we find that $r = 41^\circ$ (approx.)

The answer may also be obtained by a simple geometric construction, thus:—

Let the two media be separated (in section) by the straight line AB. From O draw the normal NON'.

Draw OC making $\angle NOC = 60^\circ$, so that CO now represents the incident ray. From OA cut off OD = 1 unit (a large unit for preference) and from OB cut off OE = 1.33 units.

Draw EF perpendicular to OB, cutting OC at F.

With centre O and radius OF describe a circle. From D draw DG perpendicular to OA, cutting the circle in G. Join OG. Then OG will represent the refracted ray, and the angle GON' will be the required angle of refraction (41°).

Proof. CO is the incident ray. To prove that OG is the refracted ray we shall have to show that $\frac{\sin NOF}{\sin N'OG} = 1.33$.

$$\begin{aligned} \text{Now } \frac{\sin NOF}{\sin N'OG} &= \frac{\sin OFE}{\sin OGD} \text{ (alt. angles)} = \frac{OE \div OD}{OF \div OG} \\ &= \frac{OE}{OD} \text{ (for } OF = OG) = 1.33. \end{aligned}$$

\therefore OG is the refracted ray, and GON' is the angle of refraction. Measuring $\angle GON'$ we find it to be 41° .

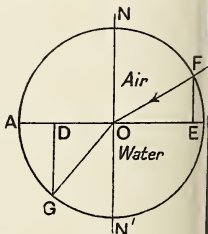


FIG. 24/8

Experiments with prisms. It is sometimes stated that when ray of light passes from a rarer medium to a denser one it is refracted towards the normal, and vice versa. Thus (Fig. 24/9) as the ray AB passes from air to the denser glass it is refracted *towards* the normal BN, while when it passes from glass to air it is refracted *away from* the normal CN'.

The 'rule' has a great many exceptions, but it applies in a number of common and important cases—air and glass, water and water, water and glass, and (a very important case) when light is passing into cold, dense air from hotter, less dense—so it is worth remembering.

Our 'dense' medium so far, whether consisting of a plate of glass or a trough of water, has always had two parallel

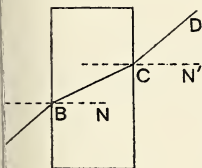


FIG. 24/9

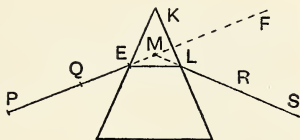


FIG. 24/10

surfaces (as in Fig. 24/9), with the result that the ray has finally emerged in a direction parallel to its original one.

A glance at Fig. 24/10 shows that this will not be the case if we use a triangular prism. In this case the ray that emerges makes a considerable angle with the incident ray (FMS in the diagram). This angle between the original and final directions of the ray is known as the *angle of deviation* (*Lat. de = from, de = a path*)—a very good term, because it really does show the angle through which the ray has been turned from its original path.

It should be noticed that when, in connection with refraction, we speak of the *angle* (or *refracting angle*) of a prism, we mean the angle between the two faces at which the light enters and leaves the prism respectively. Thus in Fig. 24/10 the refracting angle is K.

The greater the angle of the prism, the greater the angle of deviation. This is illustrated in Figs. 24/11a and 24/11b.

We shall refer back to this point when we consider the way in which a lens acts.

Tracing the path of a ray of light through a prism (either by calculation or by a graphical method) is an excellent exercise in refraction. Let us work through such an exercise.

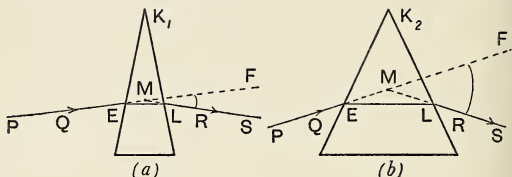


FIG. 24/11

Example 2. Draw a large diagram showing the path of a narrow parallel beam passing through an equilateral glass prism in a direction parallel to the base. Indicate on the diagram the angle of deviation (refractive index of glass = 1.5).

Suppose the beam strikes the face AB at the point D.

It passes through the prism by the path DE, parallel to the base BC.

We must now find the angle at which the ray enters the prism—i.e. the angle i which the incident ray makes with the normal NDN' .

The angle of refraction EDN' or r is 30° . For $\angle B = 60^\circ$.

$$\therefore \angle EDB = 120^\circ.$$

$$\therefore \angle EDN' = \angle EDB - \angle N'DB = 120^\circ - 90^\circ = 30^\circ.$$

$$\text{Now} \quad \frac{\sin i}{\sin r} = \mu.$$

$$\therefore \sin i = \mu \times \sin r = 1.5 \times \sin 30^\circ = 1.5 \times 0.5 = 0.75$$

$$\therefore \text{(from tables)} \quad i = 48\frac{1}{2}^\circ.$$

We therefore make an angle NDF of $48\frac{1}{2}^\circ$ and this gives us the incident ray FD . From the symmetry of the figure it can easily be shown that the refracted ray EG is obtained by making $\angle CEG = \angle BDF$.

The angle of deviation is PMG .

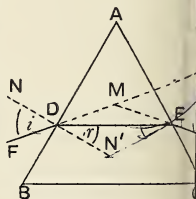


FIG. 24/12

Another solution We can also solve the problem geometrically. Draw the refracted ray DE as before. The construction is now similar to that in Example 1 (p. 268), except that there we were given the direction of the incident ray and had to find that of the refracted ray, while here it is the other way about. Try to finish the exercise for yourself.

Apparent depth of water. On p. 263 we saw that, owing to refraction, the apparent depth of water is less than its real depth. We will now try to find out *how much* less.

Suppose BD is the surface of the water and A a point at the bottom. Let AB be a ray of light striking the surface normally, and therefore continuing without change of direction along BC. AD is another ray striking the surface at D and refracted along DE. AD is so close to AB that the refracted rays BC and DE both enter an eye at CE.

Produce ED to meet BA at F. Then to the eye, A appears to be at F, and FB is the apparent depth.

We must now try to find the relation between AB (real depth) and FB (apparent depth). Draw the normal DN'. Let μ be the refractive index of water.

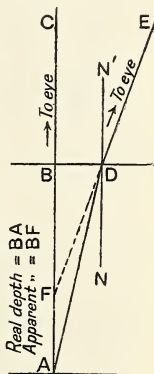


FIG. 24/13

$$\mu = \frac{\sin N'DE}{\sin NDA} = \frac{\sin BFD}{\sin BAD} \text{ (by parallels)} = \frac{\frac{BD}{DF}}{\frac{BD}{DA}} = \frac{DA}{DF}$$

But as the angle BAD is very small, $\frac{DA}{DF} = \frac{BA}{BF}$ (approx.)

$$\mu = \frac{BA}{BF} = \frac{\text{real depth}}{\text{apparent depth}}$$

This relationship would be true for any substance.

In the case of water, $\mu = \frac{4}{3}$

$$\frac{\text{real depth}}{\text{apparent depth}} = \frac{4}{3}$$

$$\text{real depth} = \text{apparent depth} \times \frac{4}{3}.$$

In Fig. 24/13 the rays from A are supposed to strike the surface normally, or very nearly so—i.e. the eye is supposed to be looking vertically downwards. It can be shown that the more obliquely the rays enter the eye, the nearer to the surface will the point F be. That is why, as you walk away from (say) a trough of drinking-water for horses, it appears to become shallower, while as you walk towards it, it seems to grow deeper.

Another method of finding μ . Since $\mu = \text{real depth} \div \text{apparent depth}$, and real depth is easily measured, it follows that if in some way we can determine the apparent depth we can find the refractive index.

The experiment is carried out as follows. We will use water as the liquid, but the method could be applied to any transparent liquid or (with a little modification) to a transparent solid such as glass.

Drop a nail into a gas jar or measuring cylinder, which is then nearly filled up with water. Looking vertically downwards into the water, the nail (A) will appear to be at A', and we have to find the position of A'.

Clamp a knitting-needle K some distance above the surface of the water. Looking downwards, an image of K can be seen, formed by reflection in the water. By moving the clamp, raise or lower K until its image coincides with A'. This is determined by the method of parallax. On moving the head from side to side, A' and the image of K should cling together.

The distance of the image of K below the surface S of the water is, of course, equal to the distance of K itself above the surface. Thus KS gives us the required apparent depth, and $\mu = AS \div KS$. By filling the jar to different depths, other answers may be obtained.

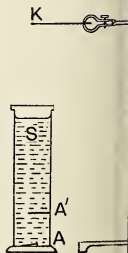


FIG. 24/14

For a transparent solid, the method is slightly modified.

Draw a straight line AB on a sheet of paper and lay a slab of glass centrally over it as shown in Fig. 24/15. Stick a pin vertically at a point on the line, and so that it touches the glass. Paste a strip of paper centrally on the upper face of the glass above AB. To avoid confusion, this strip is not shown in the figure.

Now look through the glass in the direction BA. The pin will appear to be in some such position as C'. To find C' measure

pencil point along the strip of paper until, by the method ofallax, the pencil point is found to coincide with C' . Mark this position.

Measure the distance from the 'pencil-point position' to that of the slab facing B, so getting the 'apparent depth.' The 'depth' is, of course, the full length of the slab, and by dividing the first measurement into the second, we obtain μ .

Example 3. A microscope is focused on a scratch on the bottom of a crystallising-dish. The dish is filled with liquid, and the microscope has to be raised 1.8 cm. to re-focus the scratch, and a further 3 cm. to focus a speck of dust on the surface of the liquid.

Calculate the refractive index of the liquid, deducing any formula you use in the calculation. O.C.

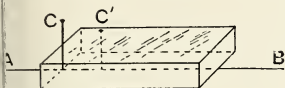


FIG. 24/15

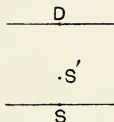


FIG. 24/16

Let S (Fig. 24/16) be the true position of the scratch, S' its apparent position, and D the speck of dust on the surface. Hence

$SD = 1.8$ cm. and $S'D = 3.6$ cm. $\therefore SD = SS' + S'D = 5.4$ cm.

$$\mu = SD \div S'D = 5.4 \div 3.6 = 1.50.$$

The latter part of the question was worked out on p. 271.

Critical Angle and Total Reflection. Suppose a ray of light travelling from air into an optically denser medium¹ such as water. The angle of refraction is less than the angle of incidence, and so for any given angle of incidence, there is always a corresponding angle of refraction. A few cases are presented in Fig. 24/17. AO , BO , CO and DO are incident rays and Oa , Ob , Oc and Od the corresponding refracted rays.

For DO the angle of incidence is 90° (or as near 90° as can be shown in a diagram). We can calculate r the corresponding angle of refraction, for since $\mu = \frac{4}{3}$ we have

$$\frac{\sin 90^\circ}{\sin r} = \frac{4}{3}$$

$$\sin r = \frac{3}{4} \times \sin 90^\circ = \frac{3}{4} \times 1 = 0.75$$

(from tables) $r = 49^\circ$ (approx.)

There is one case which will cause the ray to be refracted towards the normal. This is when the ray is incident on a medium which is often (as in this case) physically denser, but by no means always.

It appears, then, that as the angle of incidence increases from 0° to 90° , the corresponding angle of refraction increases from 0° to 49° .

Now let us imagine the direction of the light to be reverse e.g. bO is now the *incident* ray, giving rise to OB as *refracted* ray. cO is refracted as OC and dO as OD . When the angle of incidence is 49° , the angle of refraction is 90° , and cannot be any bigger. The refracted ray just skims the surface.

What will happen then with a ray such as eO , where the angle of incidence is greater than 49° ?

Such a ray cannot cross the boundary into the second medium at all. It is 'totally reflected' from it. In the diagram eO is shown reflected as OE , the angles eOa and EOa being equal in accordance with the ordinary law of reflection.

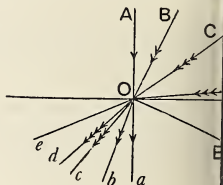


FIG. 24/17

If instead of the two media being water and air, they had been glass and air, we should have arrived at the same general results—i.e. *any* ray of light travelling in air would have been able to penetrate the glass, no matter what the angle of incidence; but a ray of light travelling in glass can penetrate the air only if the angle of incidence does not exceed a certain value. Given that in this case $\mu = 1.5$, we can easily calculate this 'certain value.' Let it be i . As it is the *biggest possible* angle of incidence, the corresponding value of r is 90° .

$$\therefore \frac{\sin i}{\sin 90^\circ} = \frac{1}{1.5}$$

$$\therefore \sin i = \frac{1}{1.5} \times \sin 90^\circ = \frac{1}{1.5} \times 1 = 0.667$$

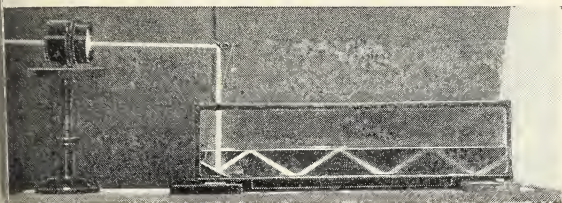
$$\therefore i = 42^\circ.$$

In the case of glass and air, then, 42° is the greatest possible angle of incidence for a ray to pass from glass into air.

¹ $\frac{1}{1.5}$, not 1.5. When we say $\mu = 1.5$, the ray is understood to be passing from air to glass, and thus $\sin i \div \sin r = 1.5$. If we reverse the direction, i becomes r and r becomes i , and so $\frac{\sin r}{\sin i} = 1.5$ or $\frac{\sin i}{\sin r} = \frac{1}{1.5}$.

angle of incidence for which a ray can penetrate the other medium. If the ray is incident at an angle greater than 42° will be totally reflected.

When a ray of light is travelling through a medium *A* towards an optically less dense medium *B*, the greatest angle of incidence for which refraction is possible is known as the critical angle for the two media.



PHOTOGRAPH OF AN EXPERIMENT SHOWING SEVERAL TOTAL REFLECTIONS AT THE SURFACE OF WATER

Note that the ray that finally issues is refracted.

From 'The Universe of Light,' by Sir William Bragg.

Unless there is some statement to the contrary, 'B' is understood to be air. Thus we say that for water the critical angle is 49° , and for glass 42° .

Notice that *two* conditions (implied in the definition just given) must be present before we can have total reflection.

- 1) The direction of the light must be towards the optically less dense medium.
- 2) The angle of incidence must exceed the critical angle.

Example 4. Calculate the critical angle for diamond, given that the index of refraction of this substance is 2.42.

Let i be the required critical angle.

Thus the corresponding angle of refraction, r , is 90° .

$$\frac{\sin i}{\sin 90^\circ} = \frac{1}{2.42}$$

at $\sin 90^\circ = 1$

$$\sin i = 1 \div 2.42 = 0.413$$

$$i = 24\frac{1}{2}^\circ.$$

Owing to this small critical angle, rays of light which enter a diamond are for the most part reflected from the next face they meet—often from a whole succession of faces—and it is this which gives a diamond its characteristic sparkle.

Example 5. Find the above angle by means of a geometric construction.

All we have to do is to construct an angle whose sine is $\frac{1}{2.42}$.

From A draw a straight line AB of unlimited length. Draw AC perp. to AB and of length 1 unit (Fig. 24/18).

With centre C and radius 2.42 units draw an arc of a circle cutting AB in D. Join CD.

Then CDA is the required angle, for $\sin CDA = CA \div CD = 1 \div 2.42$ which is the required value.

On measurement, CDA should, of course, be found to be $24\frac{1}{2}^\circ$.

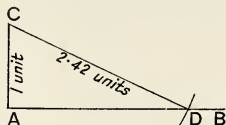


FIG. 24/18

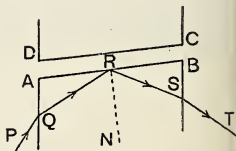


FIG. 24/19

Examples of Total Reflection. Some of these have already been discussed in *Junior Physics* (pp. 240–2), and need only be recalled briefly. ABCD (Fig. 24/19) represents (with gross exaggeration) the narrow air-space produced when glass is cracked. Rays of light such as PQR are totally reflected (the angle QRN being greater than 49° , the critical angle), and the crack is seen as a brilliant gleam.

The *mirage* often mentioned by desert travellers is the consequence of total reflection.

In contact with the hot ground, the air itself becomes hot, but at points farther and farther from the earth the air becomes cooler and cooler (and therefore denser and denser). Thus rays of light travelling downwards from an object such as A are constantly travelling from a denser medium to a rarer one, and so are refracted away from the normal (Fig. 24/20). At the same time the angles of incidence (i_1, i_2, \dots) are constantly increasing, and when the critical angle is

ached the ray begins to bend upwards (at B in Fig. 24/20). o the eye, A seems to be at A' and the sky itself gives a irage looking very like a pool, in which A (seen directly rough the cool upper layers of air) appears to be reflected. us the traveller is deluded into thinking that he is approach- g a pool of water with palm trees on the banks. The tree real enough, but not the pool.

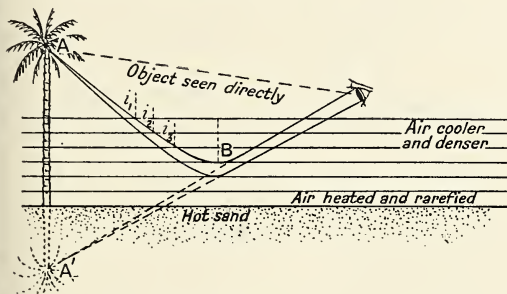


FIG. 24/20

When walking along a straight asphalted road in England on a hot day one can often see an apparent pool of water in the distance. Here the object seen as a mirage is the sky.

Fig. 24/21 shows how a glass prism of angles 90° , 45° , 45° may be used to deflect a ray of light through a right angle. This principle has a number of applications—e.g. in the periscope of a submarine, by which ships, etc., may be easily observed, though the submarine itself is submerged except for the top of its periscope.

A 'prism' mirror, for some purposes, has certain advantages over the ordinary mirror.

The latter tends to give confusing multiple images (*Junior Physics*, pp. 237–238). If the silvering is at the back, while if it is at the front it is easily damaged. The prism mirror, of course, requires no silvering, and it gives a single sharp image.

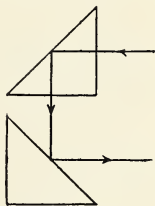


FIG. 24/21

Questions

A

1. When a lighted candle is held in front of a thick plate-glass mirror several images can be seen, one of which is much brighter than the others. Explain.

2. Draw a diagram to explain why, to an observer standing beside a swimming-bath, the water appears to be less deep than it really is.

3. Why does a straight rod partly immersed in water appear to be bent at the surface?

4. Explain *critical angle* and *total internal reflection of light* and give a labelled ray diagram to illustrate a practical application of the latter, using as an example a glass prism with angles 90° , 45° , 45° . C.W.B.*

B

1. Explain, with the help of diagrams, the terms *refractive index* and *critical angle*.

How would you determine the refractive index of glass given a glass block?

A ray of light is incident in glass on the top surface of a glass block at an angle of incidence of 60° . Assuming the refractive index of glass is $\frac{3}{2}$, find the subsequent path of the ray. Brist.

2. A ray of light makes an angle of incidence of 30° on the face of a glass prism of refracting angle 60° and refractive index 1.5. Trace the path of the ray through the prism and measure the angle of deviation.

For what other angle of incidence will the deviation have the same value? Dur.*

3. Objects frequently appear distorted when viewed through a pane of ordinary window-glass. On the other hand, the plate glass rarely produces distortion though the object may appear displaced. Explain. O.C.

4. How would you determine the refractive index of a liquid? A prism, refractive index 1.5, has a side AB of 5 cm. and angles at A and B are 45° and 60° respectively. A parallel beam of light of width 4.5 cm. falls normally on AB so that the edges of the beam are 0.25 cm. from A and B. Give a construction showing the paths of the following rays through the prism: (a) the two extreme rays, (b) the central ray. Dur.

5. Define *refractive index*.

Describe in detail how you would find (a) one of the angles of a triangular glass prism by an optical method, (b) the refractive index of the material of the prism. Lond.

6. Explain with the help of a diagram why the apparent depth of a pond is less than its true depth. Deduce the expression for the refractive index of a substance in terms of real and apparent depths. Describe an experiment to determine the refractive index of water which makes use of the relation between real and apparent depths. *Lond.*

7. A soot-covered ball when immersed in water appears brightly lished. Why?

8. A parallel beam of light falls on the surface of water at an angle of incidence of 60° . Find, by a graphical construction or otherwise, the path of the beam after it meets the water surface. The incident beam is (a) in air, (b) in water.

Refractive index of water = $4/3$.) *Lond.**

9. On a warm sunny day, dry tarmac roads frequently appear to be covered with pools of water, which dry up as the pools are approached. Why? *O.C.*

10. The following corresponding values of the angle of incidence i in air and the angle of refraction r in water were obtained:

i° .	10.0	20.0	30.0	40.0	50.0	60.0
r° .	7.5	14.8	22.0	28.75	35.15	40.5

Use these readings to obtain a straight line graph and hence calculate the refractive index of water. Draw a labelled diagram to show how you might get such a set of results yourself. You may use any special apparatus, and should give a brief description.

Camb.

1. A ray of light in air meets the surface of a rectangular glass block at an angle of incidence of 45° . It is refracted so that it meets the next one of the surfaces at right angles to the first. Find the direction on leaving this face. (Refractive index of glass = 1.5.) *J.M.B.**

2. Rays of light are emitted upwards in all directions from a point at the bottom of a swimming pool where the depth of water is 3 feet. Show that when the water is still all rays which meet the surface outside a certain circle will be totally reflected. Where is the centre of this circle and what is its radius?

Refractive index, air to water = 1.33.) *Dur.**

CHAPTER 25

REFRACTION AT CURVED SURFACES

IN ancient times—at least as early as 400 B.C.—the Greeks had either discovered for themselves or had copied from others the art of making a ‘burning-glass.’ This was shaped like Fig. 25/1a or b. You will notice that Fig. 25/1b is roughly the shape of a lentil seed, and the Latin word for lentil is *lens*. That is how such glasses came to be called lenses.

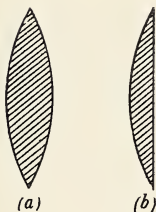


FIG. 25/1

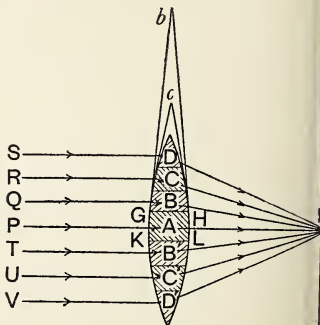


FIG. 25/2

A glass of this sort ‘burns’ because it causes rays of heat from the sun to be brought together to a point called the *focus* (*Lat. focus* = hearth—a hot place). Rays of light are affected in the same way, and it is important to notice what a lens should have such a curious effect. Suppose parallel rays of light from the sun are striking the lens as shown in Fig. 25/2, and imagine the lens to be divided into sections such as A, B, C, D.

Each of these sections is a portion of a prism, except that its sides are slightly curved instead of being plane. In Fig. 25/2 we have produced the nearly straight sides of B, making a complete prism Bb. We have also shown

complete prism Cc. We can easily see from this figure that the prism Bb is longer and sharper than Cc—it is a prism *smaller angle*.

Now, on p. 269 we saw that a prism causes a ray of light to undergo a change of direction—to be deviated—and that the greater the angle of the prism, the greater will be the deviation it produces. Hence C, a part of the prism Cc, will produce more deviation than B, which belongs to a prism of smaller angle. For the same reason, D will produce more deviation than C.

Now we can see why it is quite possible for all the rays S, Q, etc., to be brought together to the same point. To begin with, the central ray P will undergo no deviation at all, because the portion A is *not* a prism, but a plate with its opposite faces parallel. Suppose the ray Q, deviated by the 'prism' B, meets this central ray at F. Can the ray R, deviated by the 'prism' C, also pass through F? To do so, it must undergo more deviation than Q, and, of course, that is not what happens, as we have already seen. Similarly, S will need to be deviated even more, which it will be, because it is of larger angle than C.

The lenses studied in this chapter will be *spherical*—i.e. any curved surface or surfaces will be parts of spheres. With such a lens the increase of deviation needed as we work from the centre of the lens to the edge (i.e. from A to D or D') is accurately secured by the increasing 'angle of prism.' Hence the rays do not focus exactly at F. F in fact is a small circle rather than a point. The defect is known as *spherical aberration*.

Terms defined. There are certain terms constantly used in the study of lenses, and we had better get clear ideas of them at once.

To begin with the term *lens* itself. This might be defined as a transparent body bounded by two surfaces of which at least one is spherical. A **double convex** lens is one bounded by two convex surfaces. One of these is illustrated in Fig. 25/1a. If one of the surfaces is plane and the other convex (Fig. 25/1b), the lens is *plano-convex*. One surface may even be concave and the other convex (Fig. 25/3). In that case the lens is *concavo-convex*.

In all cases a convex lens is *thicker at the centre than at the edge*.

We have already seen that the curved surface (or surfaces) of a lens may form part of a sphere (or spheres), and the lens is then said to be *spherical*. The spheres may or may not be of the same radius. In Fig. 25/4 the right-hand



FIG. 25/3

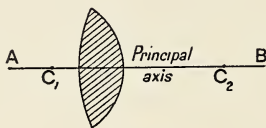


FIG. 25/4

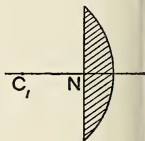


FIG. 25/5

surface forms part of a sphere of centre C_1 , and the left-hand surface has its centre at C_2 . The straight line joining the centres (AB in the figure) is known as the *principal axis* of the lens.

(In the case of a plano-convex lens, the principal axis would be the line which passes through the centre of curvature of the one face and is normal to the other— C_1N in Fig. 25/5.)

We have seen that rays of light parallel to the principal axis are brought to a focus (F in Fig. 25/2). The point is known as the *principal focus* of the lens. It can be shown that it lies on the principal axis.

Owing to spherical aberration, discussed above, it is only those rays of light passing near to the principal axis which actually pass through F. That is why a photographer 'stops down' when he wishes to get an image very sharply focused—i.e. he uses only the central part of his lens.

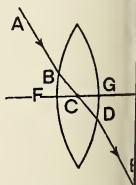


FIG. 25/6

Suppose (Fig. 25/6) a ray AB enters a lens at B and leaves at D by the path DE. If the parts of the lens in the immediate neighbourhood of B and D respectively are parallel, then for practical purposes the ray is passing through a glass plate with parallel sides, and its final direction DE will be parallel to its original direction AB (p. 265)—i.e. there will

no deviation. There will, however, be some *displacement*, cause DE, though parallel to AB, will not usually be in line with it.

Suppose BD cuts the principal axis at C.

It can be shown that *all* rays of light which emerge from a lens in a direction parallel to their original direction pass through a particular point on the axis (C in Fig. 25/6). This point is known as the *optical centre* of the lens.

If the lens is thin, the optical centre may be taken without much error as any point on that part of the principal axis which lies inside the lens—e.g. any point on FG in Fig. 25/6. In practice we usually take it as C, the middle point of FG. Again, if the lens is thin, the displacement will be negligible. Thus a ray of light passing through the optical centre of a thin lens passes on its way *as though the lens did not exist at all*. We shall see presently that we can make very great use of this fact.

The *focal length* of a lens is the distance from the optical centre to the principal focus.

Simple method of finding focal length. If we hold up a convex lens L so that it faces the sun, and hold a post-card P on the other side, then by moving P we can soon find a position at which S, an image of the sun, is focused upon it. The distance CS then gives the focal length, C being the optical centre of the lens.

Of course we cannot in practice hold lenses and post-cards and make measurements at the same time. Actually we should support L in a lens-holder and with the clamp of a retort stand. A square of three-ply painted white is better than a post-card, because it is much more rigid.

Again, we often use the laboratory window-frames, perhaps 20 ft. away, instead of the sun. As mentioned on p. 234, light from these is parallel for practical purposes, and (except on a really dull day) by using them it is easier to obtain an accurately focused image.

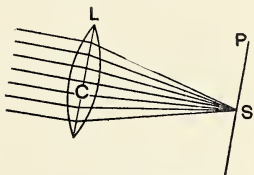


FIG. 25/7

Other methods of finding the focal length of a lens are described on pp. 286-7.

Problems solved graphically. We now have sufficient knowledge of convex lenses to work out certain problems by graphical methods, or by methods involving easy geometry (usually similar triangles).

Example 1. In the experiment just mentioned, a window-frame gives an image 1.2 in. long. The distance from the window-frame to the lens is 12 ft., and from the lens to the image 7.2 in. Find the focal length of the lens and the length of the window-frame.

The focal length of the lens is obviously 7.2 in. (position because a real image is formed at the focus).

Now consider Fig. 25/8 (not drawn to scale), in which AB is the frame, A'B' the image, and C the optical centre of the lens. DCE is the principal axis. The ray AC passing through the centre has undergone no deviation—i.e. ACA' is a straight line. Similarly BCB'. As AB is parallel to A'B' it follows that $\angle A = \angle A'$, $\angle B = \angle B'$ and so ABC and A'B'C are similar triangles.

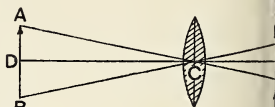


FIG. 25/8

$$\therefore \frac{AB}{A'B'} = \frac{BC}{B'C} = \frac{DC}{CE} = \frac{12 \text{ ft.}}{7.2 \text{ in.}} = \frac{144 \text{ in.}}{7.2 \text{ in.}} = 20$$

$$\therefore AB = 20 \times A'B' = 20 \times 1.2'' = 2 \text{ ft.}$$

Example 2. A small object, $\frac{1}{2}$ in. high, stands on and perpendicular to the axis of a convex lens, and a real image $1\frac{1}{2}$ in. high is formed on a screen 12 in. from the object. Make a scale diagram of the arrangement, and from it (or otherwise) find the distances of object and image from the lens, and the focal length of the lens. Lond.

Draw the axis AA', the object AB $\frac{1}{2}$ in. high, and the image A'B' $1\frac{1}{2}$ in. high (a real image formed by a convex lens is always inverted). AA' must represent 12 in., but it will be convenient to make horizontal measurements to half-scale as compared with vertical.

Join BB' cutting the axis at C. This will be the optical centre of the lens, which we can now insert in our diagram. From B draw BD parallel to

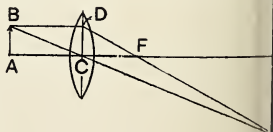


FIG. 25/9

axis, cutting the lens at D. We know that B' is the image B , and therefore the ray BD parallel to the axis must be re-
 flected in the direction DB' . Further, this refracted ray passes
 through the focus, which is therefore at F , where DB' crosses
 the axis.

By measurement we find that

distance of object from lens, $CA = 3$ in.

„ „ image „ „ $CA' = 9$ in.

focal length of lens „ „ $CF = 2\frac{1}{4}$ in.

Conjugate Foci. Going back to Fig. 25/9, suppose $A'B'$
 to be the *object*. To get the image of B' we could draw (i)
 ray $B'C$ which would proceed along the path CB without
 change of direction, (ii) the ray BFD which would be refracted
 along DB . Thus B would be the image of B' .

A pair of points such as B and B' so placed in relation to
 a lens¹ that if either of them is regarded as the object the other
 is the image, are known as **conjugate foci**.

Important lens formula. We should expect that there
 should be some purely mathematical method—independent of

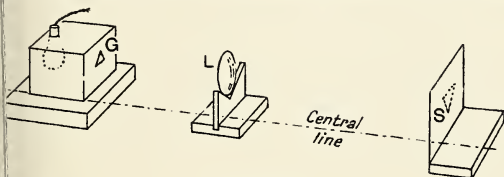


FIG. 25/10

drawing—of solving problems such as those considered
 in the examples just given. That is the case, and it can be
 shown that provided we observe the *real-is-positive* con-
 vention (p. 244) the formula already used for mirrors applies
 to lenses—i.e. $\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$.

Verification of formula. To prove experimentally that the
 formula just mentioned holds for a convex lens, we may use
 a box. A straight line is chalked on the bench, and the
 base line of the box is placed over this (Fig. 25/10). The
 top of the box consists of a piece of three-ply, out of which
 an angle has been cut to take a piece of wire gauze G . A

¹ The same applies to a mirror.

convex lens L of focal length about a foot and suitably mounted, is placed about 2 ft. from the box front, and beyond this again is a screen S .

S is moved nearer to L , or farther from it, until a sharply defined image of G is obtained. We then carefully measure the distances LG ($= u$) and LS ($= v$). The focal length f is found by the modification of the 'sun' method described on p. 283.

The work is now repeated several times (except that f need not be re-determined), altering the distance LG and then altering the position of S until the image is again sharply focused. The results may be set out thus :

Expt.	LG ($= u$)	LS ($= v$)	f	$\frac{1}{v}$	$\frac{1}{u}$	$\frac{1}{v} + \frac{1}{u}$	$\frac{1}{f}$
1							
2							
3							
etc.							

Determination of focal length. In the experiment just described, f was found by an independent method, and then from the known values of f , v and u we verified the formula. If, however, we *assume* that the formula is true, we can measure v and u as described above, and *calculate* f . In fact, the experiment gives us one of the best methods of finding the focal length of a convex lens.

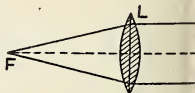


FIG. 25/11

Another method depends on the fact that if a small bright object is placed at F , the focus of a convex lens L , the rays of light which it gives off will, after passing through the lens, emerge parallel. (Imagine the rays travelling in the reverse direction, and the truth of this is seen at once.) If then a plane mirror M is placed perpendicular to the axis of the lens, these emerging rays will be reflected along their own path and will give an image at F , the image coinciding with the object.

To apply this method in practice, we arrange our ray

lens as in Fig. 25/10, but instead of a screen we have a mirror, usually placed much closer to the lens than S in Fig. 25/10. We now move the lens until a position is found in which a sharp image of the triangular gauze is formed just beside the original. The focal length is then equal to the distance GL.

Magnification. The **magnification** produced by a lens is the ratio *length of image* \div *length of object*. Suppose, for instance, that with an optical lantern a slide of 3 in. side is raised to an image of 6 ft. side, the magnification would be $12 \text{ in.} \div 3 \text{ in.} = 4$.

It can be shown by Geometry that magnification as thus defined is equal to the ratio v/u , where v and u have their usual meanings.

To verify the formula *experimentally*, we go back to the experiment illustrated in Fig. 25/10 on p. 285. After getting the image sharply focused, we use a pair of dividers to help in measuring one side of the gauze triangle, and the corresponding side of the image (not forgetting that the image is inverted). Suppose these lengths are S_1 and S_2 respectively. We find that the magnification, S_2/S_1 , is equal to v/u . We then repeat the work with other values of u and v .

The two holes of the gauze may be plugged up with bits of lathstick, the distance between the two points being taken

Concave Lenses. A characteristic of concave lenses is that they are thicker in the middle than at the edges. A concave lens on the other hand, is thinner in the middle.

Three kinds are shown in Fig. 25/12: biconcave, plano-concave and concave-concave respectively.

To understand the action of a concave lens we may suppose it to be divided into a number of prisms, A, B, B', etc., as we did in the case of the convex lens. Rays of light



FIG. 25/12

Rays . . . etc., all parallel to the principal axis, are represented as falling on these 'prism' portions (Fig. 25/13).

Comparing Fig. 25/13 with Fig. 25/2, we notice that in Fig.

25/2 each prism had its thin end pointing *away* from the centre while in Fig. 25/13 the reverse is the case. As a result, rays of light after refraction do not *converge* to a point F, but *diverge* or spread out fanwise. These diverging rays all appear to come from a point F which is thus a *virtual* focus, and so its focal length is negative.

In the previous discussion we saw that the portion C (for instance) produced a greater deviation than B, because

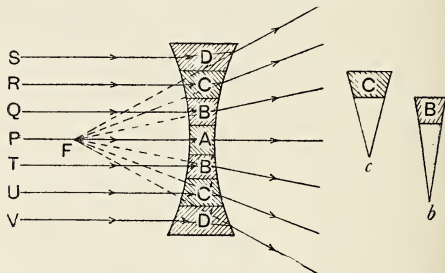


FIG. 25/13

B and C are developed into complete prisms, C will have a larger angle than B. Exactly the same reasoning applies in the present case.

Problems relating to Lenses. We will conclude this chapter by considering a few other typical problems relating to lenses. It will be noticed that they may often be solved graphically.

Example 3. An erect object, 1 in. high, is placed 8 in. in front of a convex lens, whose focal length is 6 in. Determine either by graphical construction or by calculation the position and height of the image formed by the lens. J.M.B.*

$$u = 8 \text{ in.} \quad f = 6 \text{ in.} \quad v = ?$$

The equation

$$\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

becomes

$$\frac{1}{v} + \frac{1}{8} = \frac{1}{6}$$

whence

$$v = + 24.$$

The + sign shows that the image is real. It is therefore on the side opposite to the object, **24 in.** from the lens.

$$\text{Magnification} = v/u = 24/8 = 3.$$

As the object is 1 in. high, the image will be **3 in.** high.

aphical method. It will be convenient to work to a scale of 0 horizontally, and 1 : 2 vertically.

Draw the lens, and mark its axis and focus F ; also its middle C . Put in AB ($\frac{1}{2}$ in. to represent 1 in.), BC being 0.8 in. (to represent 8 in.) and CF being 0.6 in. (to represent 6 in.) $(25/14)$.

Find the image of A in the usual way (AD parallel to the axis, join DF and produce; join AC and produce, the two lines intersecting at A'). Obtain B' by drawing $A'B'$ perpendicular to the axis.

Measurement CB' is found to be 2.4 in., representing 24 in. $A'B'$ is 1.5 in. representing 3 in.

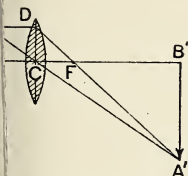


FIG. 25/14

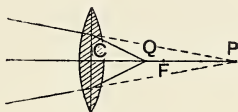


FIG. 25/15

Example 4. A beam of light is converging to a point 20 cm. from a convex lens of focal length 10 cm., the axis of the beam coinciding with the principal axis of the lens. Calculate the position of the point to which the beam converges after passing through the lens. Illustrate by means of a diagram. Lond.*

The source of light is the converging beam. The point to which this converges must be regarded as the object, and is virtual—i.e. $u = -20$ cm. The lens being convex has a positive focal length, so f (10 cm.) is positive.

$$\text{Using } \frac{1}{v} + \frac{1}{u} = \frac{1}{f}$$

$$\text{we have } \frac{1}{v} + \frac{1}{-20} = \frac{1}{10}$$

$$\text{which gives } v = +6\frac{2}{3}.$$

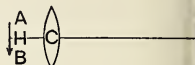
Since the image is real, the beam converging to the point Q (Fig. 25/15), where $CQ = 6\frac{2}{3}$ cm.

Example 5. A convex lens is used to project the image of an object on a screen, which is 100 cm. from the object. The object is 5 cm. square and the image 20 cm. square. Find (a) the distance of object from the lens, (b) the focal length of the lens. Camb.*

Let AB and A'B' represent the edges (5 cm. and 20 cm.) of object and image respectively (Fig. 25/16). The position of lens (centre C) has to be found.

$$\text{The magnification} = \frac{20}{5} = 4$$

$$\therefore \frac{v}{u} = 4$$



HC then must be 1/5th of the whole distance HK; i.e. 1/5th of 100 cm. = **20 cm.**

FIG. 25/16

We then have $u = 20$, $v = 80$, $f = ?$

Substituting in the formula $\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$

$$\text{we have} \quad \frac{1}{80} + \frac{1}{20} = \frac{1}{f}$$

$$\text{giving} \quad f = 16 \text{ cm.}$$

Questions

1. Distinguish between a real image and virtual image. What is meant by the focal length of a convex lens?

An object 1 cm. long is placed 40 cm. from a convex lens of focal length 10 cm. Find the size and position of the image by means of a ray diagram drawn to scale. Explain your construction.

What other information about the image does the diagram give? *Oxf.*

2. The following results were obtained in an experiment with a convex lens:

Distance of object from lens, u cm.	25.0	35.0	45.0	55.0	65.0	75.0	85.0
Distance of real image from lens, v cm.	103.2	47.0	35.9	31.1	28.9	27.3	26.2

Using your table of reciprocals, plot $\frac{1}{v}$ against $\frac{1}{u}$. Determine the focal length of the lens. Camb.

1. A convex lens forms an *erect* image which is four times the size of the object, the distance between object and image being 100 cm. Find graphically *or* by calculation the position of the lens and its focal length. *Brist.**

2. What is meant by the principal focus of a lens? Describe one accurate method of determining the focal length of a convex lens.

3. An image five times the length of the object is to be projected on a screen which is 72 cm. from the object. What kind of lens must be used, what is its focal length, and where must it be placed? *Camb.*

4. Describe how you would determine accurately the focal length of a convex lens without using a distant object.

5. A convex lens of focal length 10 cm. is used to read the graduations on a scale, and it is held so that an upright image, whose size is 5 times that of the object, is formed. How far from the scale is the lens held? *Camb.*

6. Write down a formula relating image distance (v), object distance (u), and focal length (f) for a lens. State a convention governing the signs to be given to these distances, and apply it to the formula to find the position of the image of an object placed 40 cm. in front of a *concave lens* of focal length 20 cm. Is the image (a) real or virtual, (b) erect or inverted?

7. Draw a ray diagram showing two rays from a point on the object to the eye of an observer. *Camb.*

8. Explain, with the help of a carefully drawn diagram, the formation of an image by a *concave lens*.

9. A concave lens has a focal length of 15 cm. Find the position of the image of an object placed 20 cm. in front of it. State whether the image is (a) real or virtual, (b) erect or inverted. Mention one common use of a concave lens and briefly explain it. *Camb.*

CHAPTER 26

SOME OPTICAL INSTRUMENTS

WITH a little thought we soon realise that many instruments of great practical use depend for their action on the presence of one or more lenses. There is the simple 'reading-glass', the camera, the telescope and microscope and the projector used at the 'pictures.' Among the simplest and yet the most important we have the humble pair of spectacles. There are millions of people in the world who without them would be completely cut off from the pleasures of reading, and other things, and millions who would be unable to see clearly anything at a distance.

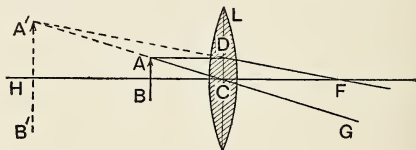


FIG. 26/1

The lens is an essential feature of many other instruments, but those mentioned will serve as a fair sample.

The Simple Microscope. This is much used as a reading-glass or pocket-lens. It consists of a convex lens, which in use must be placed at a shorter distance from the object than its own focal length. In this position it gives an enlarged, virtual image as shown in Fig. 26/1.

The eye is supposed to be just to the right of the lens, and to get the best results we move the lens about so as to bring 'A'B'' as near as possible to the eye. But there is a limit to this nearness. You know that if you are examining a small object, and you put your eye *very* close to it, you cannot 'focus' it—you cannot see it distinctly. The shortest distance at which you can see the object clearly is known as *the near point of distinct vision*. In an average case this is about 10 in.—less for young people and much more for the middle-aged and old.

We move our lens then, until $A'B'$ is at this nearest distance distinct vision.

Example 1. *A person can see clearly any object which is 10 in. further) from his eye. Make a ray diagram to show how he should best use a simple magnifying glass of focal length 4 in. and find by measurement the magnification obtained.*

We do not know the position of the object, but we know that image $A'B'$ must be 10 in. from the lens. So we begin by drawing the lens (centre C), its principal axis HCF , and the image $A'B'$ (any convenient length) 10 in. from C . We also mark F ($f = 4$ in.).

Comparing our incomplete diagram with Fig. 26/1, we see that we have to find the position of the object AB . The construction is fairly obvious. Draw $A'C$. Join $A'F$ cutting middle line of the lens in D . Then draw DA parallel to the axis, cutting $A'C$ in A . This gives us the position of the top

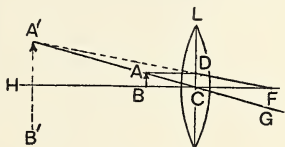


FIG. 26/2

of the object (for a ray AD , parallel to the axis, is being refracted through F) and we can easily draw AB .

By measurement, we find that the object must be **2.9 in. from lens**, and that the **magnification**, $\frac{A'B'}{AB} = 3\frac{1}{2}$.

The Camera. We have just seen that in the simple microscope the distance from object to lens must be *less* than focal length of the lens. If it is greater we get quite a different effect (cf. Fig. 25/9, p. 284). The image is now *inverted* and *real*—it can be received on a screen. Here we see, very simply, the principle of the camera, of the human eye, and of the projection lantern or optical lantern ('magic lantern').

The camera was described in some detail in our earlier book, so we shall content ourselves now with little more than a summary; but the diaphragm (Fig. 26/3b) calls for further notice. If the light is poor, the photographer will probably leave it fully open, but there is one disadvantage in this. On p. 31 we saw that an ordinary lens is subject to a defect known as *spherical aberration*—the rays striking the outer portions of the lens do not come to a focus at quite the same point as those passing through the central part. As a result, the

image formed by a fully open lens is slightly blurred. If the photographer wants a very sharply focused image, he 'stop down'—i.e. adjusts the diaphragm so that only the central part of the lens is in action. In this case, to make up for the

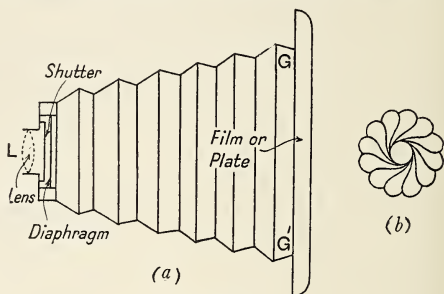


FIG. 26/3

smaller amount of light passing through the lens, he must, of course, give a longer exposure.

The Epidiascope. This instrument is often used in schools for giving an image of an object (such as a printed page, picture, map, etc.), without first preparing a slide. The object AB is very brightly illuminated, and the rays of light from it pass through a convex lens placed above it at a distance greater than the focal length. They are then reflected from a mirror silvered at the front (to avoid multiple images) and come to a focus on the screen.

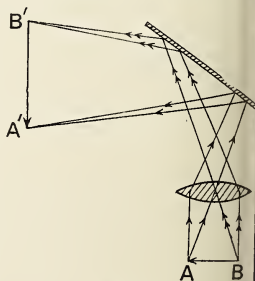


FIG. 26/4

The Eye. The structure of the eye will be clear from Fig. 26/5. Most of its parts have already been discussed (*Junior Physics*, pp. 256–7), but one or two points may now be added. The *choroid* is an inner coating which lines the outer covering

the sclerotic), except in front. The inner side of the choroid contains a dark pigment to prevent reflection of light. Attached to the choroid we have the *suspensory ligament*, which holds the crystalline lens in position. The part of the eye in front of the lens, suspensory ligament, etc., is filled with a liquid called the *aqueous humour* (A.H.) and the back part with a more gelatinous liquid known as the *vitreous humour* (V.H.). From the retina, impressions are conveyed to the brain through the *optic nerve*. The point at which the optic nerve enters the eye is known as the *blind spot*, because it is

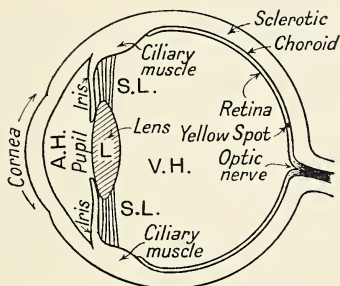


FIG. 26/5

sensitive to light. The part most sensitive to light is the *foveal spot*, near the centre of the retina.

Comparing the eye with the camera, we see that in both are (i) a lens, (ii) a diaphragm, with an aperture of variable size to admit more or less light, (iii) a screen, (iv) a shutter (the eyelid) and (v) a black coating on the inside to prevent internal reflection.

The chief difference lies in the method of focusing. In the camera, this is done by altering the distance between the lens and the screen. In the eye, it is done by altering the curvature of the lens. When we are looking at a distant object, the ciliary muscles are relaxed, and the lens has its minimum of curvature. If now we wish to see a near object, the muscles tighten, and the lens

assumes just the degree of curvature necessary to give sharp image.

This power possessed by the eye of altering the curvature of its lens is known as *accommodation*, or *accommodative power*. Towards middle life there is a marked falling off in accommodative power. A man can still see distant objects clear but cannot see to read without the help of spectacles.

Thus in Fig. 26/6 (a) rays of light from a distant object A are practically parallel, and the lens of the man's eye is able to bring them to a focus on the retina. Fig. 26/6 (b) represents a much nearer object. The rays from this are rather sharply divergent, and though the eye-lens turns them inwards as before, it is not able to overcome their original divergence sufficiently to bring them to a focus on the retina.

If, however (Fig. 26/6 (c)), a convex lens is placed in front of the eye of such focal length that the rays leave it in a parallel direction, then the eye-lens will do the rest as in Fig. 26/6 (a).

Suppose A represents print which the man wishes to read at a distance of, say, 10 in. Clearly all he needs is a convex lens of focal length 10 in.

We can put this in another way. An object which is really 10 in. away must give an image at infinity (the parallel lines to the right of the convex lens, produced backwards, meet at infinity). So $u = 10$, $v = \infty$ and f is to be found. Employing the usual formula $\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$ we have $\frac{1}{\infty} + \frac{1}{10} = \frac{1}{f}$. But $\frac{1}{\infty} = 0$ and so $f = 10$, i.e. we need a convex lens of focal length 10 in.

What we have just discussed represents an extreme case—that the man's crystalline lens has *no* accommodative power.

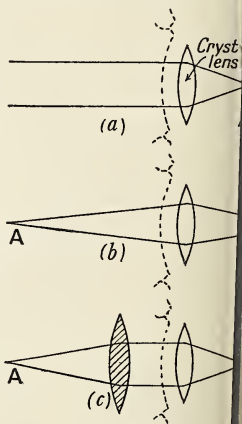


FIG. 26/6

practice it nearly always retains a certain amount. Suppose he can clearly distinguish objects at a distance of 4 ft., but not nearer (i.e. his *near point* is 4 ft.), and he wishes to be able to see things (e.g. print) at a distance of 1 ft. What length of lens will be required?

Rays from a point which is really 12 in. from the lens must seem, after passing through it, to come from a point 48 in. away—i.e. $u = 12$, $v = -48$; $\therefore \frac{1}{v} + \frac{1}{u} = \frac{1}{f}$ becomes

$\frac{1}{48} + \frac{1}{12} = \frac{1}{f}$ from which $f = 16$, so he needs a convex lens 16 in. focal length.

A defect somewhat similar to the above is *long sight*. Here the power of accommodation may be normal, but the eyeball is too thin from back to front. The result is that even when the eye is using all its power of accommodation, rays that have passed through the crystalline lens have not been brought to a focus by the time they reach the retina.

To correct the defect, a concave lens is used as already described.

Other defects of vision. Another very common defect of vision is *short-sight*, or *myopia*.

In this case, perhaps owing to some muscular weakness,

the eyeball has become somewhat bulged, and is too long from front to back. The result is that rays of light are brought to a focus in front of the retina, instead of on it (Fig. 26/a).

To put a convex lens in front of the eye would only make matters worse, for the rays of light which have passed through the crystalline lens are already too convergent, and this would make them more so. The remedy is evidently to use a concave lens (Fig. 26/b). This increases the *divergency* of the rays which strike the crystalline lens, and therefore they do not converge so quickly after they leave it.

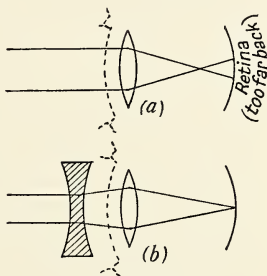


FIG. 26/7

Example 3. A short-sighted man has distinct vision only up to 4 ft. What must be the focal length of lenses which will enable him to see distant objects distinctly?

The rays of light (AB, CD) which reach the eye from a distant object will be practically parallel. We need a lens such that after the rays have passed through it, they seem to come from point P 4 ft. away (Fig. 26/8). What we need is a concave lens of focal length 4 ft. (really - 4 ft.).

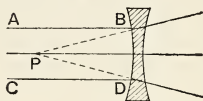


FIG. 26/8



FIG. 26/9

Or we may use our formula $\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$. $v = -4$ ft. and the object is distant we put $u = \infty$ so that $\frac{1}{u} = 0$.

Then $\frac{1}{v} + \frac{1}{u} = \frac{1}{f}$ becomes $-\frac{1}{4} + 0 = \frac{1}{f}$ whence $f = -4$ ft. as before.

It often happens, especially with elderly people, that owing to loss of accommodative power, they can see neither objects that are very distant, nor very near. In this case two pairs of spectacles are needed. You may know somebody who uses one pair of spectacles for 'the pictures' and another pair for reading. Very often the two pairs are combined in one known as 'bifocals' (Fig. 26/9) the lower part of the glass being used for reading.

The Astronomical Telescope. The best way to understand the principle of the astronomical telescope is to proceed as follows with two convex lenses, one of very short focal length say 2 in., and the other of focal length about a foot.

At one end of the laboratory pin a page of fair-sized print, say a handbill, upside down (H in Fig. 26/10). Fix the long-focus lens in a holder at L_1 , the distance HL_1 being considerable, say 10 ft. or more.



FIG. 26/10

Standing well back from L_1 you will see a real, inverted image of the handbill. Mark the position of this image by means

knitting-needle K stuck into a wooden block, using the method of parallax (p. 230).

Now place the short-focus lens L_2 about 2 in. from the knitting-needle, and putting your eye close to it, adjust its position until you get a clear large view of the needle—i.e. use the lens as a simple magnifying glass. As the real image of the print is in position K, you should now have a clear image of the handbill. In practice, you may find it necessary to re-adjust L_2 just a little. Fig. 26/11 we trace the rays by which the handbill is seen.

The first lens gives us the real, inverted image $A'B'$. The distance of this from L_2 is less than the focal length of L_2 , and so we obtain a virtual image $A''B''$ erect with respect to $A'B'$, but inverted with respect to AB .

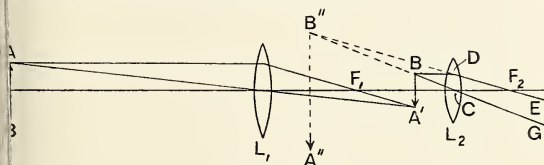


FIG. 26/11

Fig. 26/11 will serve to illustrate the action of the astronomical telescope if we understand that AB is really enormously distant, and in that case the image $A'B'$ passes through the principal focus of L_1 (cf. the 'sun' method, p. 283, for finding focal length of a lens). Again, in an astronomical observation, the distance of $A'B'$ from L_2 is equal to the focal length of L_2 . This causes emergent rays such as DE and CG to be parallel and so $A''B''$ is formed at infinity. The lens nearer to the object is known as the *object glass*, and the other lens, nearer to the eye, as the *eye piece*. From what has been said, it will be clear that the distance between the object glass and eye piece in an astronomical telescope is equal to the sum of their focal lengths.

Magnification. The definition that *magnification* = $(\text{length of image}) \div (\text{length of object})$ (p. 287) does not carry us very far because we have no direct means of finding either of these lengths. An alternative definition is adopted, viz. *magnification* = $(\text{angle subtended at the eye by image}) \div (\text{angle subtended at the eye by object})$, and it can be shown from this

that *magnification* = *focal length of object glass* \div *focal length of eye piece*.

To secure high magnification, the focal length of the object glass is often made very great.

This, of course, causes the telescope itself to be of great length since the latter has to be equal to that of the focal lengths of the object glass and eye piece added together. In 1722 James Bradley used an object glass with the enormous focal length of $212\frac{1}{4}$ ft.! The telescope had no tube, the object glass and eye piece being connected with a braced metal rod; but the practical difficulties must have been very great. These long, tubeless telescopes were known as *aerial* telescopes.

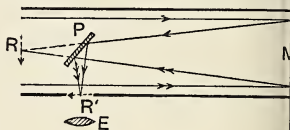


FIG. 26/12

Newton showed how to make a quite short telescope which was as effective (or nearly as effective) as these very long ones. It was known as a *reflecting telescope*, and the principle of it is illustrated in Fig. 26/12. The real image R instead



NEWTON'S REFLECTING TELESCOPE

Crown Copyright. From an exhibit in the Science Museum, South Kensington

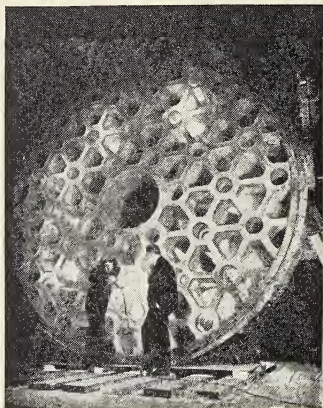
being formed by a convex lens, is formed by the concave mirror M. The rays which would give rise to R are reflected by the plane mirror P, and so the real image is formed at R'. Here it is observed through the eye piece E as in the older form of instrument. A photograph of Newton's telescope (which is still preserved) is shown in the illustration. On the side of it is the inscription 'The first reflecting telescope invented by Sr. Isaac Newton and made with his own hands the year 1671.'

Actually Newton was not, strictly speaking, the inventor of his instrument. A man named James Gregory had worked out the scheme about eight years before. But Gregory's hands were not as good as his head. He had no practical skill for himself, and could find no mechanic capable of working out his ideas. Newton was a skilled practical worker as well as a theorist.

Newton's telescope was only about a foot long, but much larger instruments have since been constructed on very similar principles. In 1745, for instance, Lord Rosse (at Parsonstown in Ireland) completed one with a mirror 6 ft. in diameter, and with a tube 58 ft. long and 7 ft. in diameter. (An Peacock once walked through it with his umbrella held up.) At the present time reflecting and refracting telescopes are both in use, some being better for one purpose and some for another. It may be of interest to give some particulars of one of the most modern instruments—a reflecting telescope which has been set up at the Mount Palomar Observatory in California. The mirror has a diameter of 6 ft., and although a 'ribbed' or honeycombed structure was adopted to reduce weight, it weighs 20 tons. Something has already been said (p. 137) of the difficulties connected with the casting of such a mass of glass, and in fact, before success was achieved, some 6 million dollars (say £1¼ million) had been spent in unsuccessful attempts extending more than five years. The mirror was finally shaped as a paraboloid (cf. p. 242). The processes of grinding and polishing were laborious in the extreme, occupying a team of fifty-two men for four years. Their leader did not desist until every square inch of the surface had been polished true within two millionths of an inch.

It is expected that with the help of this truly wonderful instrument, many hitherto unsolved problems of the universe will receive an answer.

We have said a good deal about refracting and reflecting instruments, but actually the very first telescopes made were of quite a different type. The object glass consisted of a convex lens as in the refracting telescope, but the eye piece consisted of a *concave* lens. The way in which an image was formed is shown in Fig. 26/13.



MIRROR (17 ft. diam.) OF REFLECTING
TELESCOPE AT MOUNT PALOMAR
OBSERVATORY, CALIFORNIA

By courtesy of The Corning Glass Works, U.S.A.

Rays from a distant object AB are indicated (A above the axis, B on it), passing through the convex lens O . A real inverted image $A'B'$ would be formed, but before the rays can come to a focus they strike the concave lens E . This causes them to diverge as shown, and a virtual image $A''B''$ is formed, erect in relation to the original object.

For a distant object, the distance from O to $A'B'$ would of course, be equal to F_o , the focal length of O . The distance

distance from E to A'B' is equal to its own focal length, F_e . Hence the distance between the two lenses is $F_o - F_e$. In an astronomical refracting telescope the distance would, of course, be $F_o + F_e$, so the 'Galileo' telescope, as it is called, possesses the advantage of *shortness*. A further advantage is that the image observed is *erect*. Hence we find that the Galileo telescope, or rather a pair of such, is much used for opera glasses.

The invention of the Galileo telescope is surrounded by a certain amount of mystery. 350 or 400 years ago inventors were sometimes a little coy about announcing their discoveries, for there was some danger that they would be

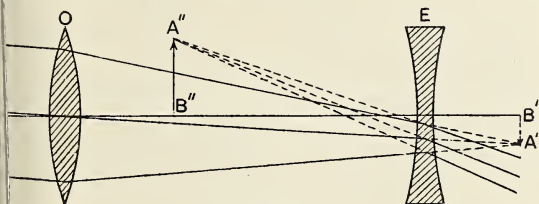


FIG. 26/13

gaged with being in league with the Evil One. One account is that the discovery was made accidentally by the children of a spectacle-maker playing with some lenses from the shop. They showed their father the remarkable views they were getting, and he followed up the matter.

That may or may not be true. It is certain, however, that about 1570 onwards there were many rumours floating about of strange magnifying effects to be obtained by the use of two lenses, and it is also certain that on October 2nd, 1608, Hans Lippershey, a maker of spectacles in Middleburg, Holland, applied for a patent. Galileo heard rumours of the invention, and very quickly discovered for himself how to make the necessary combination. Galileo, like Newton, was a very practical man as well as a great thinker, and before long he had made an instrument which brought an object more than thirty times nearer and magnified it nearly a hundred times.

Prismatic Binoculars. Going back for a moment to the astronomical telescope illustrated in Fig. 26/10, you can perhaps work out for yourself how, by using an additional convex lens, it is possible to obtain an erect image instead of an inverted one. We then have a *terrestrial telescope*. One disadvantage for common use is its length.

Prismatic binoculars, much used as field-glasses, are really a pair of astronomical telescopes, in which the length is greatly reduced by making the light pass through the tube three times. This is done by using two prisms, as shown in Fig. 26/14, arranged to give total reflection (p. 273). Notice that there are only two lenses (or lens systems), both convex, and so if it were not for the prisms the image would be inverted both vertically and laterally. The first prism, resting on its edge, corrects the vertical inversion, and the second, lying on its triangular face, corrects the lateral.

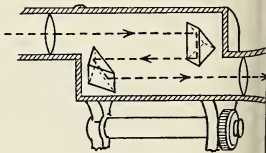


FIG. 26/14

The prismatic binocular shares with the Galileo telescope the advantage of shortness and compactness. In addition, however, it has a wide field of view, which the Galileo telescope has not.

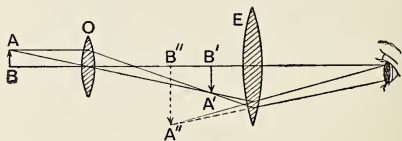


FIG. 26/15

The Compound Microscope. In this instrument the object to be examined is placed just beyond the focus of the objective glass (or *objective*), the convex lens O (Fig. 26/15). This gives us the real inverted image A'B'. E is the eye piece so placed as to be at a little less than its own focal length from A'B'. This gives us the virtual image A''B'', inverted with respect to the original object.

The image is, of course, much less bright than the original object, because the same amount of light is spread over a much greater area. To make up for this loss of brightness, the object must be very strongly illuminated.

To avoid distortion of the image, and colour effects, the object glass consists not of a simple lens, but of a system of lenses very skilfully combined, and the same applies to the eyepiece. The best modern microscopes will give magnification up to 2500 or 3000 diameters, but beyond this point the image becomes too blurred for observation.

Questions

What sort of a lens is used as a simple magnifying glass? Explain its action. Make a ray diagram to illustrate this use of a lens of 5 in. focal length, by a person who can see clearly an object 10 in. (or further) from his eye. By measurement of the diagram show that the magnification obtained is 3. *Lond.*

Describe with a ray diagram how two converging lenses may be combined to form a microscope.

A converging lens of focal length 15 cm. is used to produce a magnified image of an object with threefold magnification. Calculate the distance the object must be placed.

Give ray diagrams to illustrate the action of a camera and the eye, and point out *two* similarities and *two* differences in their optical arrangements. *C.W.B.**

The image of an object, 60 ft. away from the lens of a camera, is produced on the screen 9 in. away from the lens. In what position and through what distance must the screen be moved in order to receive a clear image of an object which is 10 ft. from the camera lens? *C.W.B.**

Describe a simple photographic camera with a single converging lens, and explain how it works. The camera of a reconnaissance aeroplane has a lens of 9 in. focal length. A photograph, enlarged 10 times from the original negative which was taken from a height of 10,000 ft., shows railway marshalling yards 27 in. long. What is the actual length of this target?

J.M.B.

The eye contains a stop, a lens system, and a sensitive photographic plate. Draw a simple labelled diagram of an eye, showing the position of these three features. (*No other details are to be included.*)

Explain, with sketches, the two defects of the eye known as *long sight* and *short sight* respectively. Describe the correction of these defects. *Camb.*

7. Why does one use convex lenses to correct long sight?

8. Explain the defect known as *short sight*, and show how it may be corrected by the use of suitable spectacles. *Oxf.**

9. The 'near point' of a short-sighted person is 4 in. in front of his eye, and he wishes to be able to read easily when a book is at a distance of 1 foot. What kind of lenses will he need, and of what focal length?

10. An elderly man cannot read a newspaper when it is held nearer than 30 in. from his eye, and he wishes to be able to read it at a distance of 10 in. What kind of lenses, and of what focal length, will be required?

11. Given two converging lenses of focal lengths 12 and 3 in. respectively, how would you arrange them to form a simple telescope? Illustrate your answer by a diagram showing how the image of a distant object is formed.

You would probably observe that when the instrument is used to view a white object the edge of the image is coloured. Account for this. *Dur.*

12. For any *one* type of telescope, comprising two lenses only, give the following information: (a) nature of eye lens (convex or concave), (b) focal length of eye lens (long or short), (c) nature of object glass (convex or concave), (d) focal length of object glass (long or short), (e) appearance of image (erect or inverted). *J.M.B.*

13. How can two lenses be arranged to form a compound microscope? State the sort of lenses used and their approximate focal lengths.

Draw a ray diagram in which at least two rays are traced through the microscope from a point on the object, not on the axis, to the observer's eye. *Camb.*

14. The 2-in. objective of a microscope is replaced by one of $\frac{1}{4}$ -in. focal length. State (*without attempting any numerical calculation*) what must be done to bring the image into focus again, and what difference the exchange makes to the appearance of the final image. *J.M.B.**

CHAPTER 27

COLOUR

our earlier studies of the subject of colour we saw that Newton had found it a source of trouble, because when an object was observed through a telescope of that period (1660) it was always seen surrounded by coloured fringes, and so the image was not sharply defined. This caused him to make experiments on colour. In one of the most important of them he made a hole about one-third of an inch in diameter in the shutter of a room, through which he admitted a beam

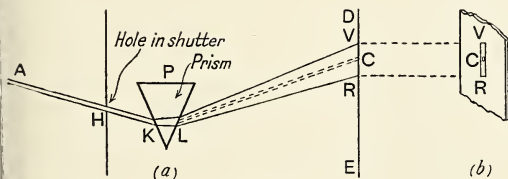


FIG. 27/1

sunlight. This passed through a prism, and was then received on a screen, where it appeared as an oblong band of colour (Fig. 27/1). The experiment is more fully described on pp. 260-1 of our earlier book. This separation of white light into its constituent colours¹ is known as *dispersion*.

As was explained on p. 200 that the different colours correspond to light of different wave-lengths, the red waves having a length of about 0.8 micron (a micron = 0.001 mm.), the violet waves of about half as much. It was pointed out too, that many wave-lengths are not represented in the visible spectrum at all, some of them being longer than the red ('infra-red'), and others shorter than the violet ('ultra-violet'). Newton suspected that the coloured beams into which the

The seven colours of the rainbow,' which are red, orange, yellow, green, blue, indigo, violet. The colours shade off from one to another; there is no sharp transition.

light had been separated could not be further broken up. However, to make sure, he used a second prism and a second screen, as shown in Fig. 27/2, which may be regarded as Fig. 27/1 continued to the right. He made a small hole in the screen so that light of a particular colour should pass through (yellow, Y, is indicated in the figure), strike the second prism and be refracted to the second screen. He did this successively with different colours, and in each case found that there was no further separation.

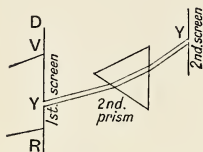


FIG. 27/2



FIG. 27/3

The spectrum obtained as in Fig. 27/1 is *impure*—i.e. the colours overlap. We can easily see how this overlapping arises. Although in the figure the yellow, for instance, the whole comes above the red, yet the yellow from the lower edge of the original sunbeam finishes below the red from the upper edge (Fig. 27/3).

By using a smaller hole we shall have less overlapping, but of course we lose a good deal of the light. To obtain a *pure* spectrum we proceed as follows.

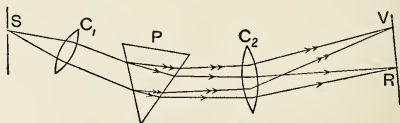


FIG. 27/4

S is a narrow slit, which may conveniently be cut in a sheet of copper. To avoid the uncertainties of sunlight, we may use artificial light—an arc-light for preference. The slit is illuminated as brightly and evenly as possible, and lens C_1 is placed at a distance from it equal to its own focal

length, so that the beam of light which emerges is parallel. This beam falls on a prism P, which is turned into the position of minimum deviation—i.e. the spectrum must be as near as possible to the patch of white light which would have appeared on the screen in the absence of the prism. As the beam emerges it falls on a second convex lens C_2 . In this way the red rays are brought to a focus at R (on a suitably placed cardboard screen), the violet at V, and the other colours at points in between.

Instead of a solid glass prism, a hollow one filled with carbon disulphide is sometimes used. This results in the production of a much wider spectrum.

The spectrum consists of a number of sharply defined images of the slit, side by side. Because the slit is narrow, the images are narrow, and there is no appreciable overlapping. If we used a really wide slit—say half an inch—we could see some of the red at one end and violet at the other. In the middle, however, all the coloured images would overlap, combining to form white.

Let us sum up the best conditions for obtaining a real, pure spectrum. They are

- (i) a narrow slit, so that the coloured images of it will not overlap.
- (ii) a prism to produce the necessary dispersion. It must be in the position of minimum deviation (strictly speaking, for the *central* portion of the spectrum).
- (iii) convex lenses suitably placed, the one to give a parallel beam, the other to bring each of the coloured beams to a focus on the screen.

It may be well at this stage to look back at our definition of Index of Refraction, given on p. 267.

We can see now that it will vary according to the colour of the light. Thus the index of refraction of water is higher for violet light than for red. Our definition, in fact, should read something like this. 'The Index of Refraction of a substance, for light of a specified wave-length, is the ratio of the sine of the angle of incidence to the sine of the angle of refraction when light of the specified wave-length passes from (or, strictly, vacuum) into the medium in question.'

Pure Virtual Spectrum. A pure *virtual* spectrum—one that cannot be received on a screen—is very easy to obtain. All we need do is to put a vertical edge of the prism parallel to the slit. On looking through the prism we shall probably see our spectrum at once. We now turn the prism until the spectrum is as near as possible to the slit—i.e. until the prism is in the position of minimum deviation. A very bright pure spectrum is observed. Fig. 27/5 shows how it is formed. For clearness, only the rays giving rise to the red and violet portions have been drawn, the red in thick lines and the violet in thin ones.

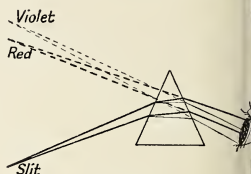


FIG. 27/5

If we look at a pin instead of the slit, the light reflected from the edge of the pin gives rise to coloured bands.

Recomposition of White Light. It appears from the foregoing experiments that white light is a mixture of lights of different colours—the colours of the spectrum. By recombining these different colours we ought to obtain white light. Newton easily showed that this could be done. Next to the first prism (Fig. 27/6) he placed a second similar one with its edge turned in the opposite direction, so that the two formed a parallel plate with a narrow gap between the two halves. He now obtained a patch of white light on his



FIG. 27/6

screen. This experiment is closely connected with the fact that if a narrow beam of light is passed through a glass plate with parallel sides, the patch of light obtained on a screen at the other side is white. We may regard our glass plate as made up of two prisms, one of them reversed with respect to the other.

Compound Colours. At this stage the reader will probably be under the impression that every colour corresponds to light of a definite wave-length, 0.8 micron for red, for instance,

is by no means the case. There are many colours—own, for example—which are not in the spectrum at all. Again, suppose we cover half of a Newton's disc (*Junior Physics*, p. 262) with red paper and the other half with green. On rotating the disc we obtain a *yellow*, which is practically the same as the yellow of the spectrum. The spectrum yellow is *pure* (i.e. it corresponds to light of a definite wavelength), while the 'red plus green' yellow is *compound*. Most of the pure colours of the spectrum can be matched by a mixture of other colours—the eye cannot tell the difference.

By means of a prism, however, we can tell easily enough. Suppose, for instance, we wish to find whether a certain colour is pure or compound. We observe a narrow strip of light through the prism to get a virtual spectrum. If the colour is compound, we shall see the component colours in the spectrum. If it is pure, we shall, of course, see only the single colour.

Absorption of Colour. Suppose we put a piece of red glass in the path of the white light that is giving rise to a spectrum—we might put it in just to the right of S in Fig. 27/4. We now find that all the spectrum disappears except the red part. Evidently, then, the red glass has absorbed all the light, except the red. Similarly, a piece of blue glass absorbs all except blue. If we put a red glass *and* a blue one in the path of the light, the spectrum is completely blotted out, as we should expect. The red glass allows only red light to pass, but this red light is absorbed by the blue glass.

We might expect that a piece of yellow glass would blot out every part of the spectrum except the yellow. Actually, it seems to make very little difference, but on closer inspection we find that it has blotted out a part in the deep blue. The light that reaches us through the glass is evidently not the pure yellow of the spectrum, but a mixture of all the colours in the spectrum except deep blue.

When you go in for photography you will have noticed that the colours seem to make too strong an impression on the plate—a blue dress or a blue flower often comes out white on the developed print. The trouble is that a photographic plate or film is unduly sensitive to the shorter wave-lengths (to the blue-violet end of the spectrum). The difficulty

is removed by putting a transparent yellow 'screen' in front of the lens. As we have already seen, this removes the deep blue.

It was at one time thought that when light passes through coloured glass, the colour was *added* to the light. We can see now that coloured glass acts not by addition, but by subtraction. That is why, if a church contains a great array of stained-glass windows, its interior will be dim.

Colour by Reflection. If we throw a spectrum on the screen as indicated in Fig. 27/4, and put a piece of red wool in various parts of it, we find that it appears red in the red part of the spectrum, but black (or nearly so) everywhere else. Red wool evidently reflects red light, but absorbs light of every other colour. Similarly, grass is green because it absorbs every colour of light except green, which it reflects.

If, however, we put a primrose petal into different parts

in the yellow part yellow, in the green part green, and so on. But in the deep blue it appears black. We are dealing with a compound colour, not a pure one, and if we look at it through a prism, we shall see a nearly complete spectrum—a series of images of the petal of every colour except

Red }
Orange } m
Yellow }
Green }
Blue }
Violet } A

FIG. 27/7

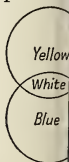


FIG. 27

blue. On the other hand, if we put some common salt into a bunsen flame, to make it give the yellow 'sodium' flame, and look at this flame through a prism, we shall see a yellow image *only*, because this particular yellow is 'pure.'

While considering this particular colour, we may recall the well-known fact that to obtain green a painter often mixes yellow and blue. The yellow pigment, like the primrose petal mentioned above, absorbs the colours at the 'blue' end of the spectrum (marked A in Fig. 27/7). The blue pigment absorbs the colours at the other end (marked B). Thus the light is absorbed except green, which is reflected.

But suppose you have two electric torches, each throwing a white circle of light on a screen. Put a yellow glass in front of one and a blue glass in front of the other. Our two circles on the screen are now yellow and blue respectively. Mix

of the torches, so that the coloured circles partly overlap (fig. 27/8). The result is not green, as you probably expected, but white!

The reason is that the yellow patch contained all the spectrum colours except blue (which was absorbed by the glass, cf. p. 311). When we added blue to this by moving the second torch, *all* the colours were assembled, giving white. We really *added* colours together, while in the case of the red and yellow paints each pigment *subtracted* something, and between them they removed everything but green.

Many other little problems on colour present themselves in daily life—and at examinations—and they are easily answered when once we have grasped the principles that we have been discussing. We can see, for instance, that if we look at a Union Jack through a piece of red glass, the red will continue to appear red, while the blue will appear black, because our red glass absorbs every colour of light except red. The white part will, of course, appear red.

A room with blue wallpaper (or with walls distempered blue) would seem rather dark if the electric light were covered with a rose-coloured shade. You can work this out for yourself.

Colours by artificial light. It should be clear by now that the colour of an object depends on two things: (i) the light that it is capable of reflecting, (ii) the light that falls upon it. A white table-cloth is capable of reflecting all colours, and in a red light would appear red, in a blue light, blue; and so on. A green dress is capable of reflecting only green, and in a red light would appear black.

The fact that the colour of an object depends partly on the light falling upon it has important consequences in the matching of colours. Suppose a lady is unwise enough to buy a hat by artificial light, to 'go with' a red dress she has at home. By daylight she will very likely find that though the dress is red, the hat has undergone a mysterious change from red to violet!

The explanation is that the proportion of red and blue in electric light is the same in electric light as in daylight. The electric light is relatively richer in red than daylight, or, what amounts to the same thing, it is poorer in blue. When the lady bought

the hat, it was reflecting red light and blue light, but not much blue, because there was not much of this colour present. On the whole, the hat seemed red. In daylight, however, the hat reflected a generous quantity of blue in addition to the red, and the red plus blue gave purple.

Similarly, many a lady has complained that a dress which was a rich blue by daylight has presented a very washed appearance in the light of a ballroom. The dress-material is evidently capable of reflecting blue freely (absorbing other colours). In the ballroom, however, the light is poor in blue and so there is not much to be reflected. It is not very like that sodium light (a pure, not compound, yellow), sometimes used for illuminating roads, will ever be used for lighting a ballroom. If it were, everybody would appear to have a yellow face, and to be dressed either in yellow or in deep mourning!

If you were asked whether you could suggest some means of altering the red:blue ratio in electric light, so as to make it nearer the ratio for daylight, you would very likely say, 'I must either cut down the red, or increase the blue. I do not see how I can very easily increase the blue, but I can cut down the red by making the light pass through a shade containing a suitable tint of blue. This blue would let through the blue that is already there, but would absorb some of the red.'

This is exactly what is done, and the result is 'daylight lighting'—a fair approximation to the balance of colours found in daylight. Of course the amount of blue in the shade must be carefully adjusted. If it is too intense, it will absorb all or nearly all the red, and the result may be worse than ever. More recently, fluorescent lighting has been used to produce the same effect.

Questions

1. Illustrate by means of a diagram the difference in the refraction, as compared with the mean value for white light, of (a) a red ray, (b) a blue ray, when passing through the same prism. Give reasons for your answer. *C.W.B.**
2. Explain why a pin when viewed through a glass prism, appears to be coloured. *Camb.**
3. Describe an experiment to show that white light can be split up into a number of coloured beams. How can these beams be recombined, and what is the result of the recombination? How would you show that any one of the coloured beams cannot be further split up? *Dur.*
4. Describe how you would set up an experiment to throw a spectrum on a screen. Draw a diagram to show the paths of the rays which give rise to the red and blue colours in the spectrum. What would be the effect on the spectrum of placing (a) a sheet of red glass, (b) a sheet of blue glass, and (c) a sheet of red glass and a sheet of blue glass in the path of the light? What would be the effect of mixing some blue and some yellow lights? Give reasons. *C.W.B.*
5. Explain why a grass lawn appears green when viewed through a certain piece of glass whether the lawn is covered with snow or not. *J.M.B.**
6. What would be the appearance of a Union Jack if viewed successively in red, blue and green light? Give reasons for your answer. *O.C.**
7. Explain why a bright blue object, viewed by sodium light, appears to be black. *Camb.*
8. A narrow strip of black paper is placed on a piece of white paper. It is viewed through a prism with the refracting edge nearest and parallel to the strip. Describe and explain what you would expect to see. *Dur.**
9. If a man stands with his back to a lamp emitting white light and looks into a thick sheet of red glass held in front of him, he sees two images of the light in the glass. Show, by diagram, how these two images are formed and state the ways in which they differ from one another. If he were to look through this sheet of glass at a white poster with red and blue letters on it, describe and explain what he would see. *Lond.*
10. Account for the appearance of a matt surface, which is (a) black, (b) white, (c) red, and (d) blue, when it is viewed in a beam of red light. *Camb.**
11. The image of a brightly illuminated object produced on a screen by a converging lens is often fringed with colour. What is the cause of this? Show clearly by a diagram what happens. *Dur.**

CHAPTER 28

SOUND—PRODUCTION AND PROPAGATION

Musical Sounds and Noises. Many sounds have a definite *pitch*; they are 'high' or 'low.' We think, for instance, of a sound produced when a violin string is bowed, or a note when the piano is struck, or the school bell is ringing. All such sounds are said to be *musical*. Many other sounds, however, have no definite pitch, as when somebody drops a heavy book on the floor, or bangs a door. On the whole, musical sounds are pleasant and noises are not, but there are plenty of exceptions. The real distinction is that in one we have a note of definite pitch, and in the other we have not. Our concern is with musical sounds, though in the course of our study we shall learn what a noise is, from a scientific point of view.

Sound produced by vibration. A little thought soon satisfies us that sound¹ is traceable to a certain kind of movement. If we pluck a violin string to one side and then release it, it gives a musical sound, and at the same time it is moving rapidly from side to side, or 'vibrating,' to use the correct term. If we put our finger on the string for a moment, and, of course, stop the vibration—and at the same time we stop the sound.

We cannot see the separate vibrations because they are too fast. By 'persistence of vision' the string seems to occupy the whole space between its extreme positions. Very similar remarks apply to the effect produced by striking a tuning-fork. In many cases the vibrating substance is a column of air, as in the case of an organ pipe, or of wind instruments such as the cornet or trombone. A circular saw gives a musical sound, which rises in pitch as the saw increases speed, and falls again as it slows down. In this case, as each tooth in succession strikes the wood, a quiver or vibration is set up in the air, and it is the rapid succession of vibrations which gives the musical note.

¹ Here and afterwards *musical* sound will be understood unless there is a statement to the contrary.

A medium necessary. In all the cases we have considered the vibrating body is at some distance from the ear which receives the sound, and between the two there is air. The question at once arises, 'Could the sound be conveyed from the vibrating body to the ear if there were no air in the space between?' It was answered somewhere about 1650 by the German Otto von Guericke of Magdeburg. Guericke had invented an air-pump, and this led him to try all sorts of 'vacuum' experiments. In one of these he put a striking block in a jar from which all the air was pumped out. The end of the striking could then not be heard. The experiment is often carried out to-day with an apparatus very similar to Guericke's, but an electric bell suspended by a thin wire is usually employed. The wires are connected to a Leclanché cell, which operates the bell. As the air is pumped out, the sound becomes fainter and fainter, until it cannot be heard at all, though the clapper can be seen striking the bell. If now we stop working the pump and open the tap to re-admit air, the sound is once more heard. It seems clear, then, that sound cannot travel through a vacuum.

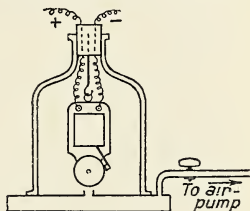


FIG. 28/1

Observations on the sun lead to the same conclusion. Using the telescope, men can see that mighty explosions are often taking place—explosions so great as to blow out pieces of the sun about the size of our earth and hurl it to a distance of perhaps a hundred thousand miles, after which it comes back. In comparison, the atomic bomb is almost nothing. Yet we never hear the faintest sound, because between us and the sun, except for a hundred miles or so from the earth's surface, there is a vacuum.

There, then, is one great difference between light and sound. Light can travel through empty space; sound cannot.

Sound can be conveyed by many other substances besides

Any substance will do, solid, liquid or gas. If you are swimming under water in the baths you can hear the shouts of people outside, sound being conveyed through the water.

Then no doubt you have played 'telephones' (many years ago of course!) with two tins of cocoa connected by stretched string. The sound is conveyed through the string. A man may hear the sound of a very distant train by putting his ear to the iron rails. Deaf people can sometimes hear the sound of a watch by putting it between their teeth, the sound in this case being conducted through the bones of the head. Many modern appliances for the deaf depend on this principle.

Pitch, Loudness and Quality. As sound is caused by vibrations, we should expect that when sounds differ from one another there will be some difference in the corresponding vibrations. We must consider, then, the ways in which vibrations differ.

BA is one prong of a tuning-fork. If we knock the end A on a table, the prong will start vibrating from one extreme position BX to another BY. Considering the point A, vibrating between X and Y, we should say that AX or AY is the *amplitude* of its vibration—i.e. the amplitude is the distance from the position of rest (A) to the extreme position reached on either side (X or Y). The fork sets up vibrations in the air—a point we shall consider more fully presently—and a section of air, PP', might vibrate between extreme positions MM' and NN'. Here again PM or PN would be the amplitude of vibration. In the case of a section of air close to the fork the amplitude will be greater than for a section farther away. Also, the sound close to the fork is louder than the sound farther away. This fact—i.e. that as the amplitude falls off, so does the loudness—strongly suggests that the loudness of a sound is connected with the amplitude of the vibration which accompanies it. We can confirm this conclusion in various ways. Thus if we tap the fork on the table gently, so that the prongs move with a vibration of small amplitude, the sound is soft. The harder we tap the fork, the greater must be the amplitude of the vibration of the prongs, and (we know by experiment) the louder is the sound.

But vibrations may differ not only in amplitude, they

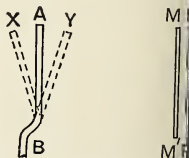


FIG. 28/2

er in *frequency*—i.e. in number per second. The prong of the tuning-fork may move to-and-fro much faster than the prong of another. In that case, of course, it makes more vibrations per second, and we say it has a *higher frequency*. The question arises: In what way would a sound produced by vibrations of high frequency differ from one produced by vibrations of a lower frequency?

A simple experiment will help us to answer the question. A long strip of steel AB, say 2 feet long, be clamped in a vise near the end B. Then push A downwards and release it. It vibrates up and down so slowly that we can see the separate vibrations. Perhaps we can almost count them. No *note*, however, is produced; the vibrations are too

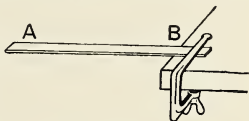


FIG. 28/3

slow. Now move AB to the right and tighten the clamp again so that the free portion of the steel is shorter. Push A downwards once more and release it. We can see that it vibrates more rapidly than before. Evidently the shorter we make the free portion of the steel, the more rapidly it vibrates.

Continue to move B to the right. A certain point is reached where we can detect a *note* of very low pitch. Then as we continue to shorten the free portion of the steel we notice we get notes of higher and higher pitch. Evidently, pitch is connected with frequency. *The greater the frequency of vibration, the higher the pitch.*¹

The same point may be proved in other ways. A circular disk of steel or cardboard has holes bored in it, equidistant from one another² and from the circumference. A piece of wood, accidentally the experiment with the steel spring gives us very good evidence that sound is caused by vibration. When the spring is short enough to produce a note, we can see the vibrations. When it is short enough to produce a note, it is reasonable to suppose that it is still vibrating, besides there is the visible 'blurred' effect produced by persistence of

the notes are bored at irregular intervals we get an effect that can only be described as *noise*—a confused sound with no definite pitch.

Thus we have a clue to the difference between a musical sound and noise. The one is produced by a movement repeated at regular intervals (i.e. by true vibration), the other by a movement repeated at irregular intervals.

glass tubing T is held by a clamp so as to be at right angles to the plane of the disc and directed at one of the holes. The glass tubing is connected by rubber tubing either to a pair of bellows or to a supply of compressed air; or if neither of these is available, one may blow with the mouth.

The disc is now rotated either by turning a suitably placed handle, or preferably by a motor. As it rotates, a puff of air passes through each time a hole comes opposite to T, and the successive puffs set up successive vibrations. By turning the wheel faster we evidently increase the frequency of the vibrations, and as we do so we observe that the note rises in pitch. When we let the wheel slow down, the pitch of the note falls. The same point is illustrated yet again by a circular saw, where the vibrations are set up by the teeth striking successively against wood, as mentioned on p. 316. Here the pitch of the note rises as the saw is being started up, remains steady while it is rotating at constant speed, and falls when it is allowed to slow down.

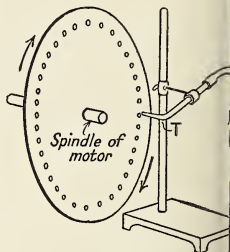


FIG. 28/4

We have now accounted for loudness and for pitch, the first being dependent on the amplitude of the vibrations, and the second on their frequency. But notes may differ in another way. Two notes, one sounded by an organ and the other by a violin, may be equally loud and of the same pitch—yet there is a difference. We say they differ in *quality*. You hear one man sing a song, and you say, 'What a lovely voice!' A little later you hear another man sing the same song, and you hope there are not many more voices. Yet the two men may have sung perfectly in tune, and with equal loudness. Again the difference is in the *quality* of the notes.

The question we have to answer is this. If two notes are of the same loudness and pitch, but of different quality, is there a difference in the corresponding vibrations? We know the difference is *not* one of amplitude or of frequency.

The explanation lies in the fact that a given tone is rarely *simple*—i.e. it does not correspond to vibrations of one frequency only, but of several. When, for instance, we sound a note on the piano, the octave and a number of other notes are produced at the same time. These extra notes are called *overtones*, the note originally sounded being called the *fundamental*. In most instruments, the frequencies of the overtones are simple multiples of that of the fundamental.

Comparing, for instance, a note of the same pitch on a violin and trumpet respectively, overtones will be present in both cases, but some which are weak in the case of the violin will be strong in the trumpet note, and vice versa; or one may be present in one case and absent in another. The overtones of higher pitch are much more pronounced in the case of the trumpet, and it is this circumstance which gives the trumpet note its ringing quality.

We can obtain some confirmation of this explanation by using two tuning-forks,¹ C and C'—i.e. upper C—each provided with its sounding-box. On sounding them together we obtain a rich, full note (C). If now we quench C' by holding the fork with the fingers, we notice an immediate change in the quality.

The *quality* of a note, then, depends on the particular overtones present, and on their strength.

Reproduction of Sound. Since all the features of a sound—pitch, loudness and quality—are definitely related to corresponding features of the vibration which produced it, it follows that if by some means we could reproduce the vibration, we should be able to reproduce the sound. We now see how this is achieved, contenting ourselves for the most part with general principles.

Suppose AB (Fig. 28/5) is a sensitive diaphragm, connected to the bent lever CDE. This makes a right-angled bend and is pivoted at P. As C moves to the right, E will move to the left, and vice versa.

The arrows represent sound-waves, say from a speaker's mouth, impinging upon the diaphragm, and causing it to vibrate. This causes corresponding vibrations of E, a needle which in which the lever terminates.

A tuning fork gives a simple note—i.e. overtones are absent.

E is in contact with a waxed plate, revolving at a uniform speed. As the plate revolves once, a circular channel produced by the needle, but the base of the channel shows a continuous wavy line owing to the vibration of E. As we have described it, E would continue to make circular channels coinciding with the first, but by means of suitable mechanism it actually traces a spiral, finishing near the centre of the plate.

Suppose we are now able to make the waxed surface very hard, and we replace the disc in the position of Fig. 28/5. We set it revolving once more. *Before*, vibrations of AB caused E to trace out a wavy path. *Now*, the hardened path, pressing against E, causes AB to repeat its original vibrations, and therefore to repeat the sound which produced them.

There is rather a lot of 'supposing' about all this, yet in broad outline it indicates how sound is actually reproduced—i.e. (i) the waves produced by the vibrating body are made to produce movements in a needle, which in turn leaves a record of its movements in wax, and (ii) the process is reversed, as already indicated.

The early records were actually traced out by an arrangement very similar to that represented in Fig. 28/5, but a much superior electrical method is now employed, rather too complicated for description here. It is, however, easy to understand how the soft path in the wax is converted into a hard one. The waxed surface is first 'metallised'—i.e. dusted over with a very fine metal powder to make it electrically conducting. By electrolysis (Chap. 34), a thin layer of copper is deposited over it, followed by one of nickel. The nickel-plated copper is now a negative or reversed copy of the original. It is separated from the waxed surface, and backed with a steel plate to strengthen it.

Discs of a black, plastic mixture looking like vulcanite are now prepared, and these are placed in turn against the negative and subjected to very high pressure. When the pressure is released and the black substance has cooled, we have a perfect reproduction of the original, but in a

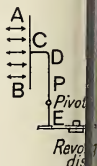


FIG. 28/5

distance instead of in soft wax. A single negative serves the preparation of some thousands of records.

How sound travels. We have already seen that sound does not travel through empty space, though it travels readily enough through air, water, iron—in fact through any material substance, whether solid, liquid or gas. *How does it travel?*

Think of a line of railway trucks, A, B, C . . . Y, Z, with the engine touching A (Fig. 28/6). The trucks are separated by the usual spring buffers.

The engine gives a sudden push towards A. What happens? The springs *a* are, of course, compressed, and A is pushed to the right. This movement causes the springs *b* to be compressed, and B in turn moves to the right, and so on, finally the springs *z* are compressed, and Z moves to the right. In short, A, B, C . . . all move in succession to the right, and as the little spaces between the buffers close up,

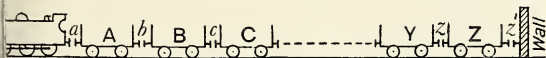


FIG. 28/6

we hear a clank, clank, clank all down the line. We say a *wave of compression* is passing along.

Now, soon after this wave of compression has started—before it reaches Z—the engine, pushed by the compressed springs *a*, will spring back to the left. The springs are now relaxed, causing A to move sharply to the left, and so on till the movement reaches Z. We hear another series of ‘clanks’ following the first series, and we say that a *wave of rarefaction* is passing down the line. In both cases (compression and rarefaction) the *wave* moves from left to right, but while in the case of the wave of compression each truck moves to the right—i.e. in the *same* direction as the wave—in the case of the wave of rarefaction each truck moves to the left—i.e. in the *opposite* direction to the wave.

Now, instead of trucks A, B, C . . . , let us imagine the *air* divided into sections A, B, C Instead of an engine moving only to the right, let it be some vibrating body, say P, prong of a tuning-fork (Fig. 28/7). P will compress A,

A will compress B, and so on, and a wave of compression passes through the air. Then P springs back, reducing the pressure on the left of A, and so A springs back, then B, and so on. Our wave of compression is followed by one of rarefaction, and it is this double wave which forms a *wave sound*. After P has gone as far as it can to the left it will of course, start moving once more to the right, giving rise to a second wave of compression. In fact, if we could see what was happening, we should see a constant succession of waves set up by the vibration of P, a compression every time it moves to the right followed by a rarefaction every time it moves to the left. Evidently one complete vibration—i.e. one complete to-and-fro movement—gives rise to one complete sound-wave, made up of a wave of compression and one of rarefaction. A 'C' fork vibrates 256 times per second. Such a fork will set up 256 complete sound-waves in one second.

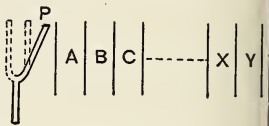


FIG. 28/7

For simplicity, the sections of air in the figure, A, B, C, are represented as all being in a straight line. Actually, of course, a body vibrating in the air produces disturbances all round it, and this sound is heard in all directions, though (usually) more in some directions than in others.

Wave-length. In our discussion of how a sound-wave travels, we have spoken of a wave of compression reaching one section of the air, while a following wave of rarefaction has reached another. We must *not* understand, however, that the intermediate sections are neither compressed nor rarefied. The transitions are gradual, not sudden. If we represent the sections of air by parallel lines, putting them closer together or farther apart to indicate the degree of compression or rarefaction, we get a state of things somewhat as shown in Fig. 28/8a.

We are much more familiar with waves on the surface of water than with sound-waves. Fig. 28/8b represents a wave passing from left to right. It is really quite different from a sound-wave, because the material substance (in this

the water) does not move in the direction of the wave, but at right angles to it. Still, it may help us if we think of our points of maximum condensation $C_1, C_2, C_3 \dots$ as being something like the crests of our water-wave, while the points $R_2, R_3 \dots$ are something like the troughs.

Now, in a succession of water-waves (Fig. 28/8b) the wave-length is the distance from crest to crest (e.g. from C_1 to C_2). Equally, it is the distance from trough to trough (e.g. R_1 to R_2), and there are many other distances that could be quoted— L_2 to L_3 , where L_2 and L_3 are said to be in the same *phase*. However, for convenience let us say the wave-length is the

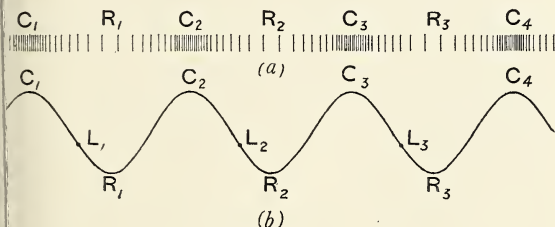


FIG. 28/8

distance from crest to crest. Similarly, the wave-length in the case of sound is the distance from one point of maximum condensation to the next (e.g. C_1 to C_2), measuring in the direction of the wave.

Next, suppose the body which is producing our sound—say the prong of a tuning-fork—is making n vibrations per second. After one second has elapsed n waves will have been sent out, and the total length of the train of waves will be $n \times l$, where l is the wave-length. But by this time—i.e. after one second has elapsed—the first condensation will have travelled a distance v , where v is the velocity of sound, and $v = n \times l$, i.e. *velocity of sound = frequency \times wave-length*.

Example. Calculate the length of the waves sent out by a C tuning fork (frequency 256), taking the velocity of sound as 1100 ft. per sec.

$$v = 1100 \text{ and } n = 256.$$

$$1100 = 256 \times l$$

$$l = 1100 \div 256 = 4.30 \text{ ft.}$$

Velocity of Sound. We have just referred to the velocity of sound as 1100 ft.—roughly one-fifth of a mile—per second. This is, of course, extremely slow in comparison with the velocity of light (186,000 miles per sec.). Let us first record a few observations which show that sound has a comparatively low velocity.

When walking along a country road you may have seen a stone-breaker at work some distance ahead. You see him hit the stone, but there is a perceptible interval before you hear the sound. You *see* the action practically at the moment it occurs, because light travels at such an immense speed. The little interval between seeing and hearing is the time taken by the sound to travel from the stone-breaker to you. When you get within 30 or 40 yd. of him the interval is so short that you cease to be able to detect it.

You can think of other examples for yourself. A very good one is the case of thunder and lightning—you see the lightning before you hear the thunder. By observing the interval you can make a rough estimate of your distance from the centre of disturbance. Suppose the interval is 10 sec. Taking the velocity of sound as one-fifth of a mile per second, this gives a distance of $10 \times \frac{1}{5} = 2$ miles.

Let us now consider how the velocity of sound has been determined.

One of the oldest methods, used about the year 1640 by the Jesuit Père Mersenne, was to get a man to fire a pistol at a known distance. Mersenne noted the interval between seeing the flash and hearing the report, and easily worked out the velocity. Thus if the distance were 4200 ft. and the interval 3 sec., the distance travelled per second would be $4200 \div 3 = 1400$ ft.

Obvious sources of error are that: (i) the time interval is so short that an error of, say, even one-tenth of a second makes a considerable difference to the result, and (ii) no allowance is made for the direction of the wind.

In 1738 several members of the French Académie des Sciences who had recognised these sources of error used cannon, 18 miles apart, at points that we will call A and B. By carrying out the experiment in both directions (A to B and B to A), and taking the average of the two results,

nated the wind effect. Further, with such a great tance the time interval was reasonably long, about 86 sec. tually, they obtained very good results.

But the velocity of sound may be found quite easily in the oratory, using simply a tuning-fork and a glass tube. The thod will be discussed later (p. 340).

Effect of temperature, pressure, etc. It was shown long ago Newton that the velocity of sound in a gas is proportional

$\sqrt{\frac{P}{D}}$ when P is the pressure and D is the density. Now

(i) If we increase the temperature of a gas we reduce its density, and so we increase the velocity of sound.

The increase is actually about 2 ft. per sec. for each degree C. rise in temperature.

ii) If we increase the pressure P , we increase the density D in the same proportion, and the two 'cancel out.' Hence *change of pressure makes no difference to the velocity.*

ii) If we increase the *humidity* we exchange some of the air for water vapour, which is less dense. Hence we get *increased velocity.*

v) For a similar reason, the velocity of sound will be increased if for air we substitute a lighter gas such as coal gas, and vice versa (for expt. illustrating this, see p. 345).

v) Wind involves bodily movement of the air, and so the velocity of sound is greater in the direction of the wind.

o summarise, *the velocity of sound is independent of the sure, but increases with the temperature (about 2 ft. per sec. ach deg. C.); it increases with the humidity, and is greater light gas than in a heavy one. It is greater in the direction e wind than against it.*

etermination of frequency. The lowest rate of vibration can give a note audible to the human ear is about 30 sec., the highest something like 30,000. How are encies determined?

ere is one method, though it would not apply very well

in the extreme cases mentioned, because of the difficulty 'matching' the note.

The instrument used is a *disc siren*. It consists of a drum-shaped metal wind-chest into which air can be blown from bellows. In the top, at equal distances from one another and lying along a circle, a number of holes are drilled, not vertically, but at an angle (see L in Fig. 28/9b, which shows the top disc in section). Resting on this fixed disc is a movable one similarly bored, but with the holes sloping in the opposite direction (U in Fig. 28/9b). When wind is blown into the chest, the air escaping through the fixed disc causes the movable one to revolve. Suppose there are 16 holes in each disc. Then in the course of one revolution the two sets of holes will coincide sixteen times, and at every coincidence there will be an escape of air from the chest. At ten revolutions per sec. (for instance) 16×10 or 160 puffs per sec. would escape from the chest, and we should get a *note* of frequency 160.

The revolution of the upper disc operates a counting mechanism as shown, and the number is recorded somewhat as on a gas meter.

Suppose we wish to find the frequency of the note produced by any given source of sound (tuning-fork, organ pipe, etc.). We set the dials to zero and (by moving a button) take them out of gear. Now we get a helper to keep sounding (say) the fork, while we operate the bellows, blowing harder or more gently until the note from the siren matches that from the fork. We now observe the time, preferably on a stop-watch, and simultaneously put the counting mechanism into gear. We keep as steadily as possible to the right note for a measured time, at the end of which we stop the mechanism.

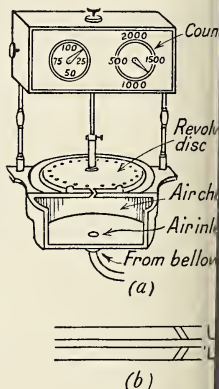


FIG. 28/9

Suppose the period was 20 seconds, and the dials indicate at in this period 360 revolutions have been made. The number of holes in the disc, we will suppose, is 16. The number of vibrations set up in the 360 revolutions is 16×360 .

But these vibrations have been produced in 20 sec.

\therefore number per sec. is $(16 \times 360) \div 20 = 288$

and so the frequency of the fork is 288.

Musical Intervals. While on the subject of frequency, it is interesting to notice that when sounds blend pleasantly into a chord there is a simple ratio between their frequencies. Thus if we strike the notes C, E, G, C' on the piano, we get the familiar chord *Doh, Me, Soh, Doh'*. We could show that these frequencies are in the simple ratio 4 : 5 : 6 : 8 by using the disc siren to help us find the frequency of each note in turn, when the results would be found to 'cancel' to the simple ratios just given. A more direct method, if the apparatus available, is to use

Savart's wheel. In one form of this we have four toothed wheels mounted on the same axle, the numbers of teeth being in the ratio 4 : 5 : 6 : 8. The actual numbers might be, for instance, 40, 50, 60,

We set the apparatus in motion (preferably with a motor) aiming at a constant, moderate speed; and then we

hold a piece of cardboard lightly, and in rapid succession, against the teeth of the four wheels. The intervals of *Doh, Soh, Doh'* are very obvious.

The notes of the familiar *diatonic scale* (*doh, ray, me . . .*), or on the piano¹ C, D, E, F, G, A, B, C) have frequencies in the ratio 24, 27, 30, 32, 36, 40, 45, 48.

C D E F G A B C

The *interval* between two notes is represented by the ratio between their frequencies, the higher one coming first. Thus

The statement is not *quite* true for the piano, but sufficiently so for all practical purposes.

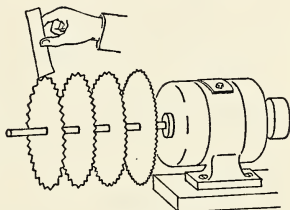


FIG. 28/10

the interval between E and C is $\frac{5}{4}$ ($\frac{30}{24} = \frac{5}{4}$), while that between upper C and lower C is $\frac{2}{1}$. This particular interval is known as an *octave*.

Reflection of a Sound-wave. It is common knowledge that when sound-waves strike against an obstruction such as the wall of a quarry, they are reflected, giving an 'echo.' Let us try to make out exactly what happens by referring back to the 'engine-and-waggon' illustration of p. 323.

Suppose the wave of compression has reached Y. This pushes to the right, compressing the spring z, which causes Z to move to the right. This compresses the spring z'. No

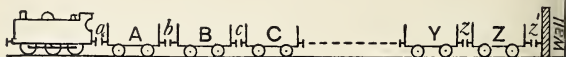


FIG. 28/11

z' cannot push the wall to the right, so being compressed, pushes against Z, making this move to the *left*. This compresses the spring z, which in turn causes Y to move to the left, and so on. Clearly the wave of compression which has travelled from A to Z is being reflected as a wave of compression travelling from Z to A. You can trace out for yourself the history of a wave of rarefaction which has travelled from A to Z, and you will find that it is reflected as a wave of rarefaction.

Leaving our illustration now and going back to the 'real thing,' we see that on reaching an obstacle, a sound-wave is reflected unchanged,¹ because the waves of compression and rarefaction of which it is made up are each reflected unchanged.

It will save trouble later if at this point we go back once more to our illustration, and consider what happens if the last truck Z is *free*.

A wave of compression travelling from A reaches Y and pushes it to the right, compressing the spring z. This pushes the truck Z, which shoots an abnormally long distance to the right because there is nothing to stop it.

¹ I.e. unchanged in character. There is always, however, some weakening as the result of absorption, and the absorption may be complete, as when sound waves strike a heavy curtain.

This sharp movement of Z to the right opens out the spring causing Y to move sharply to the right, and so on. The wave of compression which travelled from A to Z is being reflected as a wave of rarefaction. Similarly we could work the case of a wave of rarefaction travelling from A to Z—this is reflected as a wave of compression travelling from Z to A. These are the characteristics of reflection at a free end—compression reflected as rarefaction and rarefaction reflected as compression.

So far we have said nothing about the angle of reflection.

In the case of light the angle of reflection is equal to the angle of incidence

(229). When establishing this law we had to use a mirror, because the

wave-length of light is so small (a few ten-thousandths of a centimetre) that any little roughness would cause the light to be reflected in some entirely different direction.

Sound-waves are very much longer. Even the shortest sound can be heard have a wave-length of something like half an inch, and so for making experiments on reflection we do not use highly polished surfaces. A large sheet of iron or a blackboard will do very well. In Fig. 28/12 AB represents a surface. PQ and RS are tubes of tinned iron, from 6 ft. long and about 3 in. in diameter. A watch is hung at the end P of one tube, and the observer places his ear at end S of the other. A felt or baize screen XY is placed between P and S to prevent the hearer receiving sound-waves by the direct path PS.

It is found that the watch is heard best when the two tubes and the normal are all in the same plane, and when the angle of reflection is equal to the angle of incidence. Thus the laws of reflection are the same for sound as for light.

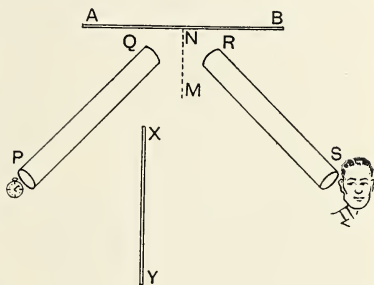
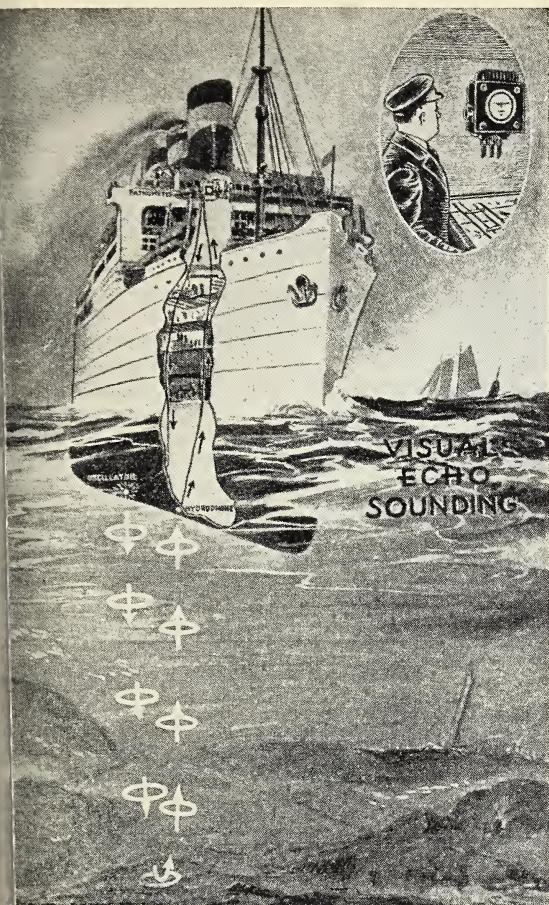


FIG. 28/12

Echoes. We observe echoes most clearly when we stand at some distance, say 100 yd., from the perpendicular face of a cliff or quarry, and give a handclap, or shout a syllable such as 'Bob.' The sound comes back to us perhaps half a second later. The sound-waves have travelled to the cliff face, and have there been reflected in the way just explained. If you are standing 275 ft. from the cliff, the sound-wave must travel there and back, i.e. *twice* 275 ft. (= 550 ft.) before you will receive the echo. Now, it takes sound one second to travel 1100 ft., so it will take half a second to travel 550

If you move close to the cliff, say to within 25 ft. of it and shout, you will not hear an echo. The to-and-fro distance is now only 50 ft., and the time taken by sound to travel this distance is only $50 \div 1100$ or $\frac{1}{22}$ nd sec. In fact, the interval between your handclap and the echo is so short that you cannot perceive it—the two sounds are merged into one. The shortest interval that most of us can detect is about $\frac{1}{10}$ th sec., and in this time sound can travel 110 ft. A distance from the cliff-face would thus be 55 ft., and an echo would be *just* perceptible. In foggy weather a ship steaming in the neighbourhood of icebergs has often obtained very important information by means of echoes. The siren gives a short blast. Suppose the echo (from the nearest iceberg) is heard in 8 sec. In this time sound would travel $1100 \times 8 = 8800$ ft. But this is the to-and-fro distance, so the distance from the iceberg is 4400 ft.—about four-fifths of a mile. The ship would, of course, proceed at greatly reduced speed, and after a minute or so would use her siren again to see whether she was moving nearer to the iceberg or farther from it. If the interval were now only 2 sec., for instance, the captain would no doubt give the order to reverse engines.

Echoes have sometimes been a great nuisance in a large lecture-hall. Every syllable heard is followed immediately by an echo, giving a confused effect, which makes it very difficult to hear what the speaker is saying. The difficulty may be got over by hanging curtains, netting, etc., over the walls, or even across the hall. These cause the sound to be absorbed instead of reflected. Sometimes a hall which gives troublesome echoes with a poor audience is quite satisfactory with a good one—sound that would have been reflected from



By courtesy of The Submarine Signal Company (London), Ltd.

THE FATHOMETER FOR VISUAL ECHO SOUNDING

Black arrows indicate electrical impulses. White rings and white arrows indicate sound waves.

an empty wooden seat is absorbed by the clothing of the person sitting on it. In the same way you must have noticed how different a room 'sounds' when it is quite empty of carpets, etc., as compared with the same room when furnished. A sound in an empty room is reflected again and again striking perhaps wall, ceiling, floor, another wall and so on.

Echoes applied to sounding. Countless ships have been stranded or broken on a rocky shore because in darkness and fog they have not known how near they were to the shore. Hence from very early times it has been customary to take soundings, usually with a piece of lead attached to a cord or wire, whenever there was uncertainty as to a ship's position. There is an early record of such proceedings in Acts, ch. v. 28.

The old method is still employed, but about 1920 a better one, in which the essential feature is an echo from the sea-bottom, came into use. The 'fathometer' method represented pictorially on p. 333, works something like this:

A loud sound is produced by the *oscillator*, a spring-driven plunger being made to strike against a steel diaphragm, the whole being suspended in water in a compartment adjacent to the ship's skin.

The sound travels to the sea bottom, and the reflected wave is received by the *hydrophone*, mounted inside the ship's skin at a distance of about 45 ft. from the oscillator.

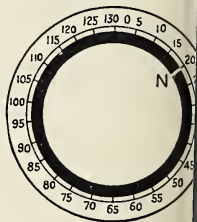


FIG. 28/13

We must now notice the *indicator*. Here we observe a fixed circle graduated in fathoms. Just inside it is a revolving circular band (shown black in Fig. 28/13), driven by a small constant-speed motor. A narrow radial slit N is cut out of this band, and under it is a neon tube which flashes red when it receives an electric impulse.

Just as the slit, in the course of its revolution, reaches the zero mark (0 in the figure) an electrical contact is set up which operates the oscillator and the plunger makes its sound.

When the hydrophone receives the echo, the vibrations are passed on to a simple two-valve amplifier, which converts

into an electrical impulse, and this lights up the neon lamp, and therefore the radial slit N. By this time N has, of course, travelled some distance from the zero position, this distance depending on the time taken by the sound to make its double journey to the sea-bottom and back again. The dial *could* be graduated in seconds, but it is obviously much more useful (knowing the speed of sound in water) to work out the depth in fathoms that corresponds to the true interval, and graduate the dial accordingly. In the figure a depth of 20 fathoms is indicated.

More than twenty soundings a minute can be made by the apparatus just described. Sometimes it is modified so as to operate a pen which gives a continuous record of the depth, rather like the barograph often seen in connection with barometers.

During recent years echo-sounding has been greatly developed. 'Ultrasonic' waves are now sent out, i.e. waves of a high frequency that the human ear cannot detect them.

(bottom of p. 327.) Since $\text{vel. of sound} = \text{frequency} \times \text{wave-length}$, and the frequency is very great, it follows that the wave-length is very short. One great advantage of using such waves is that they can be reflected by comparatively small objects—certainly by a large number of small objects, such as a shoal of fish. Since 1945, in fact, most vessels engaged in the British herring fishery have carried echosounders, which give great assistance in locating the shoals they are seeking.

Questions

How would you show (a) that sound is due to vibration, (b) that a material medium is necessary for the transmission of sound, (c) that sound travels with a definite velocity? *Dur.*

What are the physical differences between (a) a loud and a feeble sound, (b) a shrill and a deep sound?

Produce familiar evidence that sounds of all kinds travel at practically the same rate in air.

What do you understand by the 'length' of a sound-wave? If the length of a given sound-wave in air is 2 ft., what is the frequency of vibration of the body producing it? (Velocity of sound in air is 1100 ft. per sec.) *J.W.B.*

3. Describe the manner in which a sound-wave is produced and propagated through the air.

Show clearly how it is possible for musical notes of different pitch and loudness to be reproduced by a vibrating diaphragm. *Camb.*

4. Describe some form of siren. Why does its operation lead to the production of a musical sound? Explain how (a) the pitch, (b) the loudness of the note can be changed.

If the frequency of the note is 300 vibrations per sec. and the velocity of sound in air is 1100 ft. per sec., calculate the wavelength of the sound waves produced. *J.M.B.*

5. What is meant by (a) resonance, (b) beats, (c) the quality of a musical note?

How are musical intervals defined? *Camb.*

6. Explain how an echo is produced and describe an experiment which depends on echo formation, to find the velocity of sound in air.

A pistol shot is fired between two parallel high walls, 440 yd. apart, at a point 110 yd. from one wall. Find the times of arrival of the first four echoes at the firing point and give a diagram showing how the echoes are produced. (The velocity of sound in air is 1100 ft. per sec.) *Lond.*

7. A stationary ship sounds its hooter when near a cliff, and an echo is heard 3 sec. later. Assuming that the velocity of sound is 1100 ft. per sec., how far is the ship from the cliff? *J.M.B.*

8. Explain the formation of an echo.

A man stood facing a high brick wall 275 ft. away from him, clapping regularly. He noticed that when he was just clapping ten times in 5 sec., each clap coincided with the echo of the previous one. What value did he deduce for the velocity of sound in air? *Camb.**

9. Explain fully:

(a) How 'fingering' a violin string alters the note produced

(b) Why the sound of your voice appears to be very loud when you are speaking in an empty room. *J.M.B.*

10. Explain the conditions required for an echo to be heard. A ship sounds its siren when it is 3850 ft. from a vertical cliff. Taking the velocity of sound in air as 1100 ft. per sec., calculate how long it is before the echo is heard. At the moment of sounding its siren a small depth-charge is exploded in the sea near the ship and its echo from the undersea part of the same cliff is heard by hydrophones on the ship after $1\frac{1}{2}$ sec. What is the velocity of sound in water? Upon what factors does the velocity of sound in air depend? *Lond.*

CHAPTER 29

RESONANCE

p. 327 it was mentioned that we could find the velocity of sound in the laboratory by means of very simple apparatus. Before we can understand the experiment, however, we shall have to consider the subject of *resonance*.

Suppose we take two forks of the same pitch (e.g. both C, frequency 256), each mounted on its sounding-box. We set one of them strongly, and after a second or two quench the vibration by pinching the prongs. It will be found that the other fork is now sounding.

Or remove the front of a piano so as to expose the strings. Press down the loud pedal and sing any single note into the instrument. After you have stopped, the piano will be giving out the note you have sung.

To understand what is happening let us think of a person swinging on a swing. Give him a gentle push, and later, another, another, timing your pushes to agree with the natural period of oscillation of the swing. In this way the pushes are added together, so to speak, and the swing is soon describing a big arc. When you are 'working up' on a swing, you follow the movement of your body in the same sort of way.

Sometimes when walking across a plank over a brook you may have found that it moves up and down rather alarmingly. The period of your step happens to coincide with the natural period of oscillation of the plank. The remedy is obvious. Soldiers are sometimes instructed to break step when crossing a wobbly bridge to prevent a dangerous degree of oscillation.

After an air-raid it was often thought advisable to remove the walls of houses, churches, etc., which had been left standing. One method was to pass a stout cable round the building, the other end being attached to an engine (Fig. 29/1). The latter would then give a succession of tugs, timing them so as to coincide with the period in which the wall was swaying. After a few minutes the wall would come down.

Now we can see what happens when you sing into the piano. One of the strings has a period of vibration cor-

responding to the note you sing, and the vibrations that supply act upon it in exactly the same way as the timpani impulses acted on the swing, or the plank, or wall. The phenomenon is known as *resonance*, and the string is said to be set into *resonant vibration*. A similar explanation applies of course, in the case of the two tuning-forks.

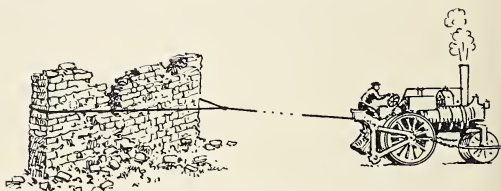


FIG. 29/1

Notice, too, that a tuning-fork does not give out much sound until we make its lower end touch a table or box. The fork itself can set only a small quantity of air vibrating. When it touches the box, however, the large wooden surface is set into 'sympathetic vibration' (as we often call vibration produced by resonance), and this in turn sets a large quantity of air vibrating. Thus we obtain a much louder sound.

We can set a column of air into resonant vibration. Take a tube about a yard long and an inch wide, closed at one end. Hold a vibrating tuning-fork over the open end. Probably the sound is very faint. Now pour in a small quantity of water and try again with the fork. Repeat the process again and again. You will find that when the water has reached a certain level the column of air resonates to the fork with a pronounced *hum* (Fig. 29/2).

In practice it will be found much easier to regulate the length of OC by using the

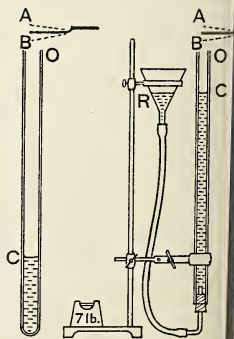


FIG. 29/2

FIG. 29/3

arrangement shown in Fig. 29/3 and raising or lowering the reservoir R.

Let us look a little more closely into the matter. In the figure, for clearness, only one prong of the fork is shown. In the position A it is just beginning to move downwards, sending a wave of compression down the tube. This will be reflected at the closed end C—the water surface—as a compression (p. 330), and when it reaches the open end O it will be reflected as a rarefaction (p. 331).

Now, if at this moment the prong has reached the position B (i.e. just about to begin its upward journey), it will itself be sending a rarefaction down the tube, which will reinforce the rarefaction produced by reflection.

This rarefaction will be reflected at C as a rarefaction, and when reaching the open end O will be reflected as a compression. At this time the fork will have completed another half-vibration, and will be back in the position A—just ready to send another compression down the tube to reinforce the compression produced by reflection. And so on. For resonance, the condition is obviously this. The time taken for a sound-wave to travel *twice* the length of the tube must correspond to the time required by the fork to make half a vibration, or we may say, while the fork is making one vibration, sound must travel *four times* the length of the tube. But the distance sound travels while the fork is making one vibration is the *wave-length* of the sound produced. For wave-length is the distance between two successive compressions such as C_2 and C_1 (Fig. 28/8a on p. 325), and a compression is produced with each vibration.

So if l is the length of the closed tube, which resounds to the fork, the wave-length of the sound is $4l$.

By lowering the side tube, let us increase the length OC. Resonance ceases, as we should expect. If, however, we continue to increase OC until it is *three times* as long as it was in the first resonance position, we secure resonance once more. Why?

Suppose t is the period of vibration of the prong. It moved from A to B in $\frac{1}{2}t$ seconds, and because we had increased the length OC, the return wave had not reached the open end of the tube. But t seconds later—i.e. $1\frac{1}{2}t$ seconds after

leaving A—the prong is in position B again. The return wave will have just arrived—it has had three times as long a journey as in the first case considered, but it has had three times as long ($1\frac{1}{2}t$ seconds instead of $\frac{1}{2}t$) in which to do so, and so we shall obtain resonance. Similarly, we should obtain it if the length OC were made $5l, 7l \dots$ where l is the shortest length that gives resonance.

Now let us see how we can use our tube and fork to find the velocity of sound in air. We must know the frequency of the fork (we saw on p. 327 how this could be found, but we may make use of the number stamped on the fork). Suppose it is 160, and that the shortest value of OC for which the tube will resound is 20.6 in.¹

The fork makes one vibration in $1/160$ th sec., or half a vibration in $1/320$ th sec., and in this time sound travels twice the length of the air column—i.e. 2×20.6 or 41.2 in. Our problem then is simply this. If sound travels 41.2 in. in $1/320$ th sec., how far will it travel in a second?

Clearly it will travel 41.2×320 in., which works out to 1099 ft.

By a similar experiment we could find the velocity of sound in any other gas, by simply filling the tube with the gas in question. If the gas is lighter than air (e.g. coal gas) it will obviously be necessary to have the closed end of the tube upwards.

In the calculation just given the frequency of the fork is known, and we were able to find the velocity of sound. Conversely, if the velocity of sound is assumed, we can find the frequency of the fork.

Example. A tuning-fork is sounded in a tube of 1 in. diameter, the length of the air column being adjustable by means of water. The shortest air column for which resonance can be obtained is 12.8 in. Calculate the frequency of the fork.

The velocity of sound at the temperature of the room (15°C.) may be taken as 1120 ft. per sec.

¹ Actually it is found that reflection does not take place exactly at the open end of the tube, but a little way above it; and in accurate work a length equal to $0.3d$ would be added to our 20.6 in., d being the diameter of the tube.

In calculations, if d is given, it is an indication that $0.3d$ is to be added as in the example which follows almost immediately.

Let n be the required frequency.

Then the time of one vibration is $\frac{1}{n}$ sec. and of half a vibration $\frac{1}{2n}$ sec.

In $\frac{1}{2n}$ sec. the sound travels twice the length of the tube, which allowing for the 'end correction') we must take as $12.8 + 0.3 = 13.1$ in.; so in $\frac{1}{2n}$ sec. sound travels 26.2 in.

But also, in $\frac{1}{2n}$ sec. sound travels $\frac{1120}{2n}$ ft.

$$\text{or} \quad \frac{1120 \times 12}{2n} \text{ in.}$$

$$\therefore \quad \frac{1120 \times 12}{2n} = 26.2$$

$$\therefore \quad 2n \times 26.2 = 1120 \times 12.$$

$$\text{which gives} \quad n = 256.$$

Resonance from open tube. Suppose we hold a tuning-fork over an *open* tube. The prong is at A, beginning to move upwards and sending a compression down the tube (Fig. 4). At the open end O' the condensation is reflected as rarefaction, which on reaching O is reflected again as a condensation. For resonance, the prong must now be in position A once more—i.e. the length of the tube must be such that sound can make the to-and-fro journey in the time of one vibration. In the case of the closed tube it had to make the to-and-fro journey in the time of *half* a vibration. Evidently, then, the shortest¹ open tube that will produce resonance with a given fork is twice as long as the shortest closed tube.

The air inside the tube. When a tube, open or closed, is sounding, there are evidently waves travelling down the tube at the same time that reflected waves are travelling in the opposite direction. The air is said to be in a condition of *stationary undulation*. Without attempting a detailed explanation of what happens, we may give a brief statement of the consequences of this double movement.

'Shortest,' because obviously we could obtain resonance by making the tube twice, three times, etc., as long, the reasoning being similar to that already given for a closed tube.

At certain points $N_1, N_2, N_3 \dots$, spaced at equal intervals (Fig. 29/5a), the air is *still*. That is because at these points it would be moving in one direction owing to the direct wave and in the opposite direction owing to the reflected one, the net result being that it does not move at all. These points of no movement are called *nodes*.

Half-way between these there are points $A_1, A_2, A_3 \dots$ where the air has maximum movement. These points are

known as *antinodes*. There is always an antinode at an open end, and always a node at a closed end. The distance from a node to an antinode is equal to one-quarter of a wave-length.

A very simple case occurs when we are finding the velocity of sound by measuring the shortest closed tube for which resonance occurs (p. 339).

There the air column AN



FIG. 29/4

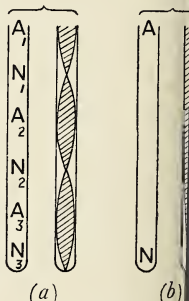


FIG. 29/5

(Fig. 29/5b) has only one node (at N) and one antinode (at A). Thus the wave-length $= 4 \times AN$, and the velocity of sound is found by multiplying this length by the frequency of the fork.

A common method of representing stationary undulations is by a pair of lines opening out to a maximum at an antinode and intersecting at a node. The method will be clear from the figure.

Note produced by blowing. Most boys know that by blowing across the mouth of a test-tube a note of definite musical pitch is produced (though science masters are often singularly unappreciative). Let us modify the experiment a little.

Take a piece of tubing about 9 in. \times $\frac{3}{4}$ in. If before cutting it off it has been slightly drawn out at one end (Fig. 29/6), so much the better. Fit it with a rubber stopper mounted on a wire. This must be close-fitting, but not so tight that it cannot be drawn in and out.

Now sound an 'upper C' fork above P and adjust the upper until resonance is obtained. Next, simply *blow* across P.

The same note is obtained. This result is found to be general—i.e. a tube or pipe gives out the same note when blown across as it does when used to give resonance with a tuning fork.

Actually the blowing *is* producing resonance, though at first it is not obvious how the blast of air can act like the tuning-fork previously discussed.

Suppose it is acting in the direction AB (Fig. 29/7). This produces a compression which travels down the tube, and, after being reflected from the bottom, travels up again. When it reaches the mouth of the tube

- (i) It pushes the layer of air at AB into the position AB'—i.e. it changes the direction of the blast. This movement causes a *rarefaction* to start down the tube, just as would a similar movement of the prong of a tuning fork.
- (ii) It is itself reflected as a rarefaction, which combines with the rarefaction produced in (i).



FIG. 29/7

29/6

Thus the blast acts for practical purposes like the prong of a tuning-fork. The difference is that the latter is so strong as to have an *independent* movement, while the movement of the blast of air is controlled by the returning compressions and rarefactions.

The *frequency* with which the air at AB vibrates—and therefore the note given out—depends, of course, on the length of the air column, which can be altered by adjusting the cork, or (using a test-tube) by pouring in more or less water. Using this principle, take a rack of, say, twelve test-tubes (Fig. 29/8) and make the middle eight give the notes *ray, me . . . doh'*. The other test-tubes will enable you to go two notes above *doh'* and two below *doh*. With

a little practice you will be able to give remarkable rendering of 'God Save the King,' 'Poor Old Joe,' etc.

Let us carry out a few more experiments with the tube Fig. 29/6. (We shall get better results if we do our blowing through a piece of tubing of about $\frac{1}{2}$ in. diameter, somewhat flattened at the end, Fig. 29/9.)

Close the end Q with your thumb, and blow across P. No repeat, but with Q open. The second note is an octave higher than the first—i.e. it has twice the frequency. This agrees with the conclusion reached on p. 341, where we saw

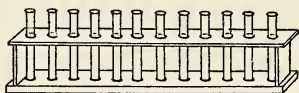


FIG. 29/8



FIG. 29/9

that the 'closed tube' note has a wave-length of $4l$, while the open tube gave a note of wave-length $2l$.

Now frequency \times wave-length = velocity of sound (p. 321)

$$\therefore \text{frequency} = \frac{1100 \text{ ft. per sec.}}{\text{wave-length (in feet)}};$$

so the 'open tube' note, having only half the wave-length the other, will have twice the frequency—i.e. it will be an octave higher.

Overtones. Using the open tube, blow first gently, and then harder. The first note you get—the lowest it can give—is called the *fundamental*, say *doh*. On blowing harder you get the octave *doh'*, and blowing harder still you may get *soh* and then even *doh''*. If n is the frequency of the fundamental, these higher notes have frequencies $2n$ (*doh'*), $3n$ (*soh'*) and $4n$ (*doh''*). And if we could blow hard enough we should obtain notes of frequencies $5n$, $6n$, etc. Such notes are the *overtones* mentioned on p. 321.

Now repeat the work with the closed pipe. Calling the lowest note—i.e. the fundamental—*doh*, we find that instead of our series being *doh*, *doh'*, *soh'*, *doh''*, *me''* . . . as with the open pipe, we obtain *doh*, *soh'*, *me''* . . . (probably you

to be able to blow hard enough to obtain me''). If n is the frequency of the fundamental, these overtones have frequencies $3n, 5n, \dots$

Any note whose frequency is a whole number times n is known as a harmonic. Thus with an open pipe it is possible to obtain *all* the harmonics as overtones, but with a closed pipe only the odd harmonics.

Organ-pipes. Fig. 29/10 shows the essential features of an organ-pipe. It may be made of wood or metal. Air from the *wind-chest* rushes through the narrow opening O and strikes the lip L. Essentially this is like blowing across the mouth of a tube. The pipe produces its fundamental note, and sounding along with this, but not so strongly, are the overtones. The pipe may be open or closed, and in the latter case, of course, the overtones are limited to the odd harmonics. The quality of the note depends on the overtones present and their relative strength. An open pipe gives a purer note than a closed one because of the greater number of overtones present.

Other wind instruments. The principle that musical notes may be produced by vibrations set up in a column of air is, of course, the essential feature of wind instruments. We will glance briefly at a few of them.

In some, such as the fife and flute, the player blows across a column of air, which in some instruments—e.g. the flute, 'tin whistle'—is open, in others—e.g. an ordinary whistle¹—closed. To obtain a higher note the player shortens the column of air by opening a hole in the side of it, a process equivalent to cutting the column at that point.



FIG. 29/10

Mr. J. Goodier (Science Masters' Book, *Physics*, Series II) shows how a whistle may be used to show that the velocity of sound in coal-gas is greater than in air. 'A whistle is attached by a fairly long piece of rubber tubing to a gas tap. The gas is turned on and the whistle sounds its normal note, due to the air in the tube being driven through. The note then rises sharply as the coal gas comes along.' The higher note, of course, indicates greater frequencies, and so sound waves must be making their little to-and-fro journeys more quickly when the whistle is filled with coal gas, than when it is filled with air.)

In the clarinet the air column is nearly closed by a flexible cane reed (Fig. 29/11). As it vibrates the column is alternately opened and closed, a little puff of air being admitted to the pipe each time it opens, and producing resonance, much as in the case of the organ-pipe. As the reed is not very rigid its rate of vibration is determined chiefly by the rate at which pulses travel up and down the pipe. Partly, too, its rate is controlled by the player's lips, and to some slight degree by its own rigidity.

A clarinet player can 'tune up' to some extent. He does it by slightly pulling out, or pushing closer together, the sections of his instrument. He thus gets a longer or shorter air column, and consequently a lower or higher pitch. You may have noticed, too, that an inexperienced player on the clarinet can produce overtones all too easily.

The oboe, bassoon and saxophone have mouthpieces very similar to that of the clarinet.

In the bugle, cornet and a number of other instruments, the vibrating lips of the performer take the place of a reed, the rate of vibration being controlled by the rate at which pulses travel through the column of air. In the cornet the player is constantly altering the length of this by means of keys. In the bugle there are no keys, and the player makes all his calls by means of overtones.

The mouth-organ strictly has no place here, because it does not work on the 'column-of-air' principle. The reed is rigid enough to have an independent rate of vibration—it is not controlled by the pulses in a column of air. The air supplied by the player escapes in a succession of little puffs, the frequency, and therefore the pitch, depending on the rate of vibration

of the reed (Fig. 29/12). In fact the action somewhat resembles that of the disc siren. The 'rigid-reed' principle is also used in the accordion and harmonium. An organ may also have 'reed pipes,' as distinct from 'flue pipes.'



FIG. 29/11



FIG. 29/12

Questions

1. Explain carefully what is meant by resonance. Describe how you would arrange a state of resonance between a column of air and a given tuning-fork, and explain how you could use the arrangement to enable you to determine the velocity of sound in the air, the frequency of the fork being known. *J.M.B.*

2. A vibrating tuning-fork labelled 400 was held close over the top of a glass tube, closed at the lower end, while water was poured in. When the level of the water was 21 cm. below the top of the tube the note of the tuning-fork swelled out, but died away again when more water was added. Explain this and draw an important numerical conclusion from the experiment.

Lond.

3. How does the velocity of sound depend on (a) the atmospheric pressure, (b) the temperature?

4. If you were provided with a test-tube rack and a dozen similar test-tubes, how would you use these to make a simple musical instrument? If the test-tubes were tuned to give consecutive notes of an ordinary musical scale (eight notes to the octave), and if the third test-tube from one end were tuned to middle C, what would be the frequencies of the fifth, seventh and tenth tubes? What relation would exist between the wavelengths of the notes emitted by the third and tenth tubes? middle C = 256. *Lond.*

5. A tuning-fork of frequency 330 per sec. is held vibrating over the open mouth of a tube about 1 metre deep, while water is slowly sent into the tube through an inlet at the bottom. A loud note is heard when the length of the air column above the water is 74 cm., and as more water is introduced the sound dies away, returning at full strength when the length of the air column is reduced to 24 cm. Explain why this happens, and calculate the velocity of sound in air from these observations. *J.M.B.*

CHAPTER 30

STRETCHED STRINGS

A GREAT many of the musical instruments we see around us consist essentially of stretched strings; the violin, piano and harp are examples.

Stringed instruments, most commonly the harp, are often mentioned in the Bible. There is a list of instruments seven times repeated in the third chapter of Daniel, 'the sound of the cornet, flute, harp, sackbut, psaltery, dulcimer, and all kinds of musick.' Of these six, at least three, and possibly four, were stringed instruments. But there are very many earlier records of stringed instruments. The illustration is a reproduction of an ancient

Egyptian drawing showing several varieties of harp. Notice the two on the right — remarkably like bows, but with a number of strings instead of the usual one. It has been suggested that the first musical sound to arrest the fancy of early man was the twang of his own bow-strings. Perhaps at the end of the



By courtesy of 'Encyc.

ANCIENT EGYPTIAN HARPS

day three or four of these primitive hunters would sit in the cave or hut obtaining different notes by twanging the different bows. Then one bright fellow thought of stretching another string across, so obtaining a two-note instrument — and the step from that to the crude form of harp seen in the old drawing would not be a very great one.

At a very early date, too, man must have gained some rough idea of the laws of strings which we shall study in the present chapter. He would find, for instance, that if he increased the tension on his bow-string, he obtained a note of higher pitch, and that this also happened if he exchanged

thick bow-string for a thinner one. Later his knowledge would become more exact, and the Greek Pythagoras, who lived about 500 B.C., is known to have discovered that if a string is divided into two lengths, then the simpler the ratio between these lengths, the more perfect is the harmony between the notes given out by the two parts.

Transverse Waves. Let us begin by recalling the kind of wave produced when you throw a small pebble into a calm pond. Just where the pebble enters, the water rocks up and down (strictly speaking, first down, then up). This movement is passed on to the next portion (a small circular belt), then to the next, and so on. The *disturbance* or *wave* moves very rapidly towards the edge of the pond. The *water* moves only up and down—not sideways—as you can see by watching a bit of floating wood or cork. Such a wave—i.e. one in which the material, in this case water, is moving at right angles to the direction of the wave—is said to be *transverse*.

What is the kind of wave that passes along a stretched string or wire when you pull it to one end and then release it. The wave travels *along* the wire.

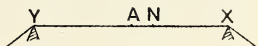


FIG. 30/1

Each portion of the material, in succession, moves *at right angles* to the length.

Suppose our wire is stretched between two fixed points Y and X (e.g. a violin string, stretched between the tuning peg and the bridge). When the wave reaches Y it is *reflected*, and a moment afterwards it is reflected again from X, and so on. Thus waves are passing *in both directions* along XY, as a few pages back we were considering sound-waves moving in both directions along a tube or pipe.

It may happen that a point A which is just being pushed forwards because of the right-to-left wave is also being pushed forwards by the left-to-right wave. As a result we get a specially high crest, while next to it there will be a specially deep trough. On the other hand, there will be points, such as B, which are stationary, because the right-to-left wave and the left-to-right wave are trying to move them in opposite directions, and simply cancel one another out.

Without going fully into detail, it is important to notice

the general result. Fig. 30/2 shows the actual movement of a portion of the wire as the two waves each advance a distance of one wave-length, the one to the left, the other to the right. We notice that certain points, $N_1, N_2 \dots$, spaced at regular intervals, are stationary. These points are called *nodes*. Midway between these are points of maximum movement (A_1, A_2 , etc.), where the crests and troughs are strongly marked. These are *antinodes*. It is found that the distance between two successive nodes (or between two successive

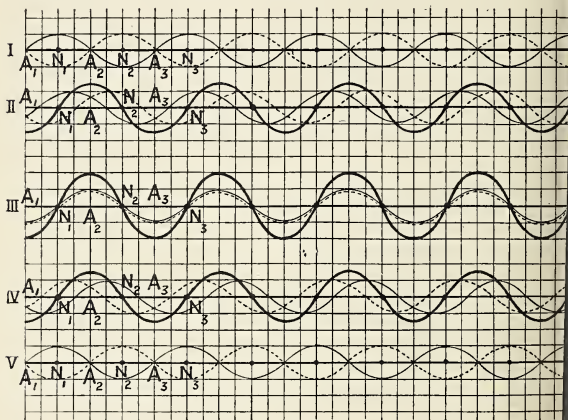


FIG. 30/2

antinodes) is equal to half a wave-length. It follows that the distance from a node to the nearest antinode is a quarter of a wave-length.

The wave-motion just discussed is known as *stationary undulation*, because the nodes and antinodes, for instance, are fixed in position. Do not confuse transverse waves with stationary undulation. A transverse wave is *progressive*—the crests and troughs are continually advancing, as we see when the pebble has just been thrown into the pond. Stationary undulation is the effect produced by the 'interference' of two transverse waves, one direct, the other

ted. You sometimes see a good illustration near a pier promenade, where the reflected waves meet the waves still coming in. A small boat is rocked up and down very violently when it happens to be at an antinode.

The presence of nodes and antinodes in a vibrating string

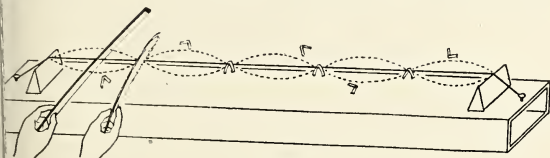


FIG. 30/3

wire may be shown by an interesting experiment first carried out by two Oxford scientists in 1676. A number of paper riders are placed over the wire. The latter is then made to vibrate by means of a bow, and is simultaneously touched with a feather at an exact fraction of its length (one-fifth in FIG. 30/3) from one end. The point touched, being unable

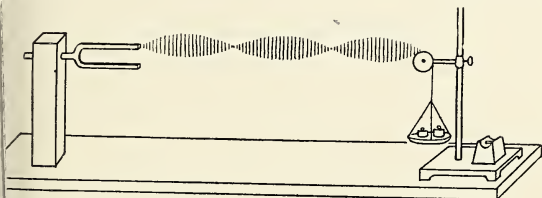


FIG. 30/4

move, becomes a node. Other nodes are formed at equal intervals, shown by the fact that at these points the paper riders keep their places, but are thrown off everywhere else.

In the experiment just described the vibrations, of course, die away, because we have made no arrangements for reinforcing the waves as they are reflected from one end or the other. This reinforcement is secured in what is known as Helmholtz's experiment. The string is attached at one end to a tuning-fork of frequency 100 or less, while the other

passes over a pulley and is stretched by a weight (Fig. 30). The fork is made to vibrate by first squeezing the prongs and then releasing them. The pulley is then moved away until a position is found in which the string vibrates in one loop. On reducing the stretching force to one-quarter (keeping pulley and fork in the same positions) the string can be found to vibrate in two loops, and if the force is reduced to one-ninth it will vibrate in three.

Laws of strings. In our earlier studies (*Junior Physics*, pp. 283-6) we saw that the frequency of vibration of a string depends on (i) its length, (ii) the force with which it is stretched, and (iii) the mass of unit length. We must now go into the subject more exactly.

In carrying out our experiments we shall use a *sonometer*. In its usual form this consists of two strings stretched across bridges, which are mounted on a long, rectangular wooden box (the box, of course, acting as a sounding-board). The best results are obtained if the box is hung against a wall (Fig. 30/5), the lower portion being made to project an inch or so.

An iron hook, on to which weights can be slotted, is firmly attached to the lower end of the string.



FIG. 30

Frequency and length. Sufficient weights are placed on F to make AB when plucked give a clear note, rather low note. CD, which is simply a 'reference string' is tuned to the same note by turning the tuning peg (not shown). Call this note *Doh*.

A movable bridge G^1 is now placed under AB, and moved until the length AG when plucked gives *Doh'*, the octave above the reference note sounded by CD. It will be found that $AG = \frac{1}{2}AB$. Thus when the length is halved, frequency is doubled.

Let us experiment a little further. We shall assume

¹ G must, of course, be of sufficient height to press against the string—but only just; if it pushes AB up appreciably it will increase the tension, and this should be constant.

relative frequencies of *doh*, *me*, *soh*, *doh'* are as 4 : 5 : 6 : 8. This can be proved to be the case by finding the four frequencies by means of the disc siren—the experiment is described on p. 328—or by the ‘Savart’s wheel’ experiment, (329.)

By moving the bridge, find the length that gives *me*, and then the length that gives *soh*. We may set out our results like this.

Note	Doh	Me	Soh	Doh'
frequency	$4n^*$	$5n$	$6n$	$8n$
length (cm.)	100	80	66·7	50
product, frequency \times length	$400n$	$400n$	$400n$	$400n$

Thus if the reference string is sounding C (frequency 256), $n = 64$, the actual value of n does not matter.

The law now stands out clearly enough. If a string is kept under constant tension, the product of frequency times length will be a constant; or, as it is more usually expressed, *if a string is kept under constant tension, its frequency of vibration will be inversely proportional to its length.*

An alternative method is to use several tuning-forks of known frequency, say 256, 320, 384, 512. The tension on the string is adjusted so as to make it give the lowest note required (in this case frequency 256), and then, keeping the tension constant, the length is altered so as to make the string give in succession the notes agreeing with those of the other forks. In this case we should have no need of a reference string. The results are set out as above, but of course n does not appear. It is worth noting that the law just established gives us an alternative method of comparing the frequencies of two tuning-forks. We adjust the bridge until the wire is in tune with the first fork (frequency n_1), and measure the length, l_1 , between the bridges. We do the same for the second fork (frequency

If the length is now l_2 , then $\frac{n_1}{n_2} = \frac{l_2}{l_1}$.

Frequency and Tension. We begin as before, AB giving the reference note and CD being in tune with it. This time, however, we shall need to know the weight of the iron support. Suppose

this is 0.8 kilos, and that the weight slotted on to it is 2 kilos total 2.8.

We now put on extra weights until AB is sounding octave. Suppose the attached weights are now 10.4 kilos. Adding 0.8 for the weight of the support, we have a total $10.4 + 0.8 = 11.2$. Now 11.2 is just *four times* 2.8. To to double the frequency we have had to increase the tension fourfold.

It looks as though the frequency is proportional to *square root* of the stretching force. To test this, we find what tension is required to make the string sound *me*, and afterwards *soh*. With careful work, such results as the following would be obtained.

Note	Doh	Me	Soh	Doh'
Frequency .	$4n$	$5n$	$6n$	$8n$
Stretching force .	$0.8 + 2.0$ $= 2.8$	$0.8 + 3.6$ $= 4.4$	$0.8 + 5.5$ $= 6.3$	$0.8 + 11.2$ $= 12.0$
$\frac{\text{Frequency}}{\sqrt{\text{Stretching force}}}$	$\frac{4n}{\sqrt{2.8}}$ $= 2.39n$	$\frac{5n}{\sqrt{4.4}}$ $= 2.39n$	$\frac{6n}{\sqrt{6.3}}$ $= 2.39n$	$\frac{8n}{\sqrt{12.0}}$ $= 2.39n$

Thus frequency $\div \sqrt{\text{stretching force}}$ gives a constant quantity (2.39n in our experiment) and so frequency is *proportional* to $\sqrt{\text{stretching force}}$. The law may be formulated thus: *For a string of constant length, the frequency of vibration is proportional to the square root of the stretching force.*

As in the case of the previous law (connecting frequency with length), we may use tuning-forks instead of a stretched string. If we used the same forks as before, the 'frequency' line of our results would read 256, 320, 384, 512, and in the last line we should have

$$\frac{256}{\sqrt{2.8}} = 153. \quad \frac{320}{\sqrt{4.4}} = 153. \quad \frac{384}{\sqrt{6.3}} = 153. \quad \frac{512}{\sqrt{11.2}} = 153.$$

Frequency and Mass per Unit Length. By experiments somewhat similar to those already described it can be shown that the frequency of vibration of a wire is *inversely* proportional to (i) its diameter, (ii) the square root of the density.

the material of which it is made. It can be shown (by very difficult mathematics) that both these results are included in the statement *the frequency of vibration of a string is inversely proportional to the square root of the mass per unit length*.

Mathematical expression. If frequency is denoted by n , length by l , stretching force by t and mass per unit length by m , the laws stated on pp. 353-5 may be expressed thus:—

$$(i) \ n \propto \frac{1}{l}, \quad (ii) \ n \propto \sqrt{t} \quad \text{and} \quad (iii) \ n \propto \frac{1}{\sqrt{m}}$$

Three of which are combined in the statement

$$n \propto \frac{1}{l} \sqrt{\frac{t}{m}}$$

$$n = \frac{1}{l} \sqrt{\frac{t}{m}} \times k$$

where k is some constant.

If l is measured in centimetres, t in dynes (= wt. in grams $\times 981$) and m in grams per centimetre of length, it can be shown that $k = \frac{1}{2}$ and so $n = \frac{1}{2l} \sqrt{\frac{t}{m}}$.

Example 1. A steel string vibrates between two bridges 100 cm., the stretching force applied to it being 10 kilograms. One length of the wire weighs 0.24 gm. Calculate the frequency of a tuning-fork which is in unison with this string.

We have to find n , given that $l = 100$, t (10,000 gm.) = 10×981 dynes and $m = \frac{0.24}{100} = 0.0024$ gm. per cm.

In the formula we have

$$n = \frac{1}{200} \sqrt{\frac{10,000 \times 981}{0.0024}}$$

This is best worked out by logarithms, and we get $n = 320$.

3. This exercise suggests a means of determining the frequency of a tuning-fork, organ-pipe, etc. We stretch a meter string by means of a weight, and then by moving bridges make it give the same note as the fork. We measure l (distance between bridges) in cm., and observe t (stretching force in gm. $\times 981$). We get m by weighing a known length of the wire and dividing by the number of cm.; n is then calculated from the formula.

Example 2. A stretched steel string emits a note of frequency 256 when vibrating under a certain tension. When the tension is increased by 5 kilograms the note emitted by the wire is of frequency 384. What was the original tension of the wire. Lond.*

Let x = original tension in kilograms.

Since frequency \propto square root of tension, then

$$\frac{\text{1st frequency}}{\text{2nd frequency}} = \frac{\sqrt{\text{1st stretching force}}}{\sqrt{\text{2nd stretching force}}}$$

$$\therefore \frac{256}{384} = \sqrt{\frac{x}{x+5}} \quad \text{or} \quad \frac{2}{3} = \sqrt{\frac{x}{x+5}}$$

Squaring,
$$\frac{4}{9} = \frac{x}{x+5}$$

Multiplying across, $9x = 4x + 20$

$\therefore x = 4 \text{ kilograms.}$

Or we could have used the full formula.

In the first case, $256 = \frac{1}{2l} \sqrt{\frac{1000x \times 981}{m}}$

„ „ 2nd „ „ $384 = \frac{1}{2l} \sqrt{\frac{1000(x+5) \times 981}{m}}$

$\therefore \frac{256}{384} = \sqrt{\frac{x}{x+5}}$, giving $x = 4 \text{ kilograms}$ as before.

Production of harmonics. Pluck a sonometer wire AB in the middle and observe the pitch of the note. Now touch lightly with your finger at its middle point M, pluck it at a point one-fourth of its length from one end and immediately afterwards remove your finger. The original note, say *doh*, gives place to the octave *doh*'.

At first the string was vibrating as a whole, or, as we sometimes say, 'in one loop' (dotted lines, Fig. 30/6a).

The second time we kept the middle point M at rest touching it, and the string then vibrated in two loops (solid lines, same figure). M, in fact, was made a node. AB vibrating as if it were two separate strings, AM and MB, if we were gripping it at M with a pair of forceps.

And since each of the 'separate strings' was half the length of AB, it vibrated with double the frequency, and so gave the octave.

By touching the wire at N, one-third of its length from

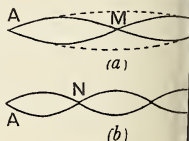


FIG. 30/6

l, and then plucking it (immediately afterwards removing finger from N), we make it vibrate in three loops, N being node (Fig. 30/6b). The note given out is now *soh'*, with three times the frequency of the original note *doh*.

The lowest note given out by the wire vibrating as a whole is called the *fundamental* or *first harmonic* of the wire. The note given out when it is vibrating in two loops is called the second harmonic, or first overtone. When vibrating in three loops we get the third harmonic, or second overtone, and so on.

In the experiments described above only one sound has been produced at a time, either the fundamental or one of the harmonics. In practice—e.g. when a violin is being played—the fundamental and one or more of the harmonics are sounded *at the same time*, and the quality of the note depends on which harmonics are sounding, and on their relative strength.

Application to Instruments. It is interesting, and quite worth noticing, to notice how the laws of strings—dependence of frequency on length, tension, etc.—are illustrated by well-known musical instruments. In the violin, for instance, the string, used for the lowest notes, is the thickest, and the string, for the highest, is the thinnest. Further, to increase mass per unit length, the G string is wound round with silver-faced wire. The violinist tunes up by using the pegs (one for each string) to increase or reduce the tension. Although he has only four strings, he can obtain any note he wants by altering the effective length of the string, which he does by pressing it against the finger-board with his finger.

In the piano many of these features recur. The tuner adjusts the pitch by altering the tension of the strings with the tuning-hammer. For mechanical reasons, however, the tuner takes care that there shall be no great difference of tension in different parts of the instrument, and he secures high notes and low ones by making the strings for the former short and thin, while those for the latter are long and thick. The lowest are heavily wound with copper wire—two layers of it—to increase the mass per unit length. To secure extra length for the low-note strings, the piano is *over-strung*—i.e. the low-note strings pass diagonally

over some of those of higher pitch. Even in the ordinary upright piano this diagonal arrangement often secures effective length of 4 ft. or so, while that for the strings of highest pitch is scarcely as many inches.

Questions

1. Describe how you would verify the relation between length and frequency of vibration of a stretched string.

State the effect on the frequency if (a) the length and tension were both multiplied by four, (b) the thickness of the string were increased, the length and tension being kept constant. *J.M.B.*

2. The lengths of the vibrating segments of a monochord (sonometer) when the wire was in tune with forks of the following frequencies were as follows:

Frequency	100	150	200	250	300
Length (cm.).	150.0	100.0	75.2	60.2	50.0

State the law to which you would expect these readings to conform and verify that they do so, preferably by a suitable graph.

What do you understand by the term wave-length? Calculate its value for a note given by a closed pipe, 16 ft. long, which is vibrating in its fundamental. *Camb.*

3. The tension in a wire 45 cm. long is 25 kg.wt., and it vibrates with a frequency of 320 vibrations per sec. It is made to sound in unison with a musical note of frequency 360 vibrations per sec. by altering either (a) the tension, or (b) the length. Find the required tension and length respectively. *Dur.**

4. Under a tension of 4 kilograms a certain wire 60 cm. long gives a note of frequency 256, when vibrating. How could you make it give a note of frequency 512, (i) altering only the length, (ii) altering only the tension?

ANSWERS TO NUMERICAL EXERCISES

Answers to the odd-numbered questions only have been given

Chapter 13, p. 135

A

1. 212, 122, -40, 32, 59.

B

1. 3 cm.

Chapter 14, p. 143

1. 298° C. 3. 182° C. 5. 1.0051 c.c.; 8.889 gm. per c.c.
100 cm.

Chapter 15, p. 149

1. 1.024 gm. 5. 29.8 cm. 7. 22 cm.

Chapter 16, p. 161

A

1. 273, 288, 0, 373. 3. (i) 546 c.c.; (ii) 746 c.c.

B

162° C. 3. 233.2. 5. 87.1 cm. 7. 73.8° C.
1076 mm. 11. 760 c.c. 13. 91.0 or 91.1 c.c. 15. 36° C.

Chapter 17, p. 172

A

380. 3. (i) 0.855 cal.; (ii) 22.8 cal.

B

18.3 gm.; 0.45. 3. 0.67. 5. 707. 7. 2.89 : 1.

Chapter 18, p. 182

A

(i) 100; (ii) 800; (iii) 1000; (iv) 5400. 3. 76.4° C.

B

17.7° C. 3. 132.5 gm. 5. 3.03 cal. per gm. 7. 0.132 gm.
1.25 gm.

Chapter 21, p. 219

A

3. 10.

B

1. 44.2. 3. 1. 5. 0.070° C.**Chapter 22, p. 247**

B

7. 25 cm. in front; 1 cm.

Chapter 23, p. 260

1. 4 ft. 3. (1) $1\frac{1}{4}$ ft.-candles (2) 2.83 ft. 5. 0.6 ft.
 7. 2.83 ft. from screen.

Chapter 24, p. 278

B

11. Reflected at $\angle 62^{\circ}$ with normal.**Chapter 25, p. 290**

1. $\frac{1}{3}$ cm. long; $13\frac{1}{3}$ cm. from lens, on opp. side.
 3. $33\frac{1}{3}$ cm. from object; $44\frac{1}{3}$ cm. 5. 8 cm.
 7. $8\frac{1}{4}$ cm. from lens, same side as object (a) virtual; (b) erect

Chapter 26, p. 305

5. 3000 ft. 9. -6 in.

Chapter 28, p. 335

7. 1650 ft.

Chapter 29, p. 347

5. 33,000 cm./sec.

Chapter 30, p. 358

1. (a) Halved. 3. (a) 31.64 kg.wt.; (b) 40 cm.

PART III

MAGNETISM AND ELECTRICITY

CHAPTER 31

MAGNETISM. GENERAL PHENOMENA

Early history. There is evidence that man has been interested in magnets for much more than 2000 years. These early magnets were found in the earth, and consisted of pieces of mineral now known to have been an oxide of iron with the formula Fe_3O_4 .

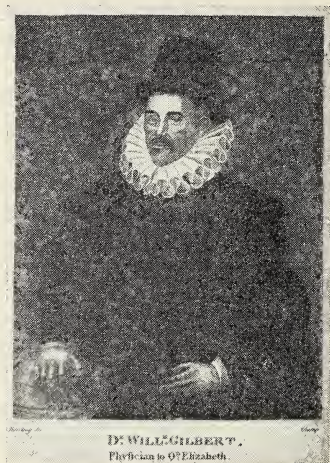
The first fact to be discovered was the obvious one that they would attract pieces of iron. Much later, perhaps in the 12th century, it was found that a magnet if suitably suspended, or if placed on a piece of wood floating on water, always pointed in the same direction. But along with these important facts, many absurd ideas were entertained. It was believed, for instance, that 'if a loadstone be anointed with garlic, or if a diamond be near, it does not attract iron'; and that 'if pickled in the salt of a sucking fish, there is power to pick up gold which has fallen into the deepest wells.' Such a stone might, of course, get into the wrong hands. It was 'of use to thieves . . . being a stone born as it were to theft'; it could even open bars and locks.

There were other dangers, too. There were actually mountains in the north of such great powers of attraction that ships are built with wooden pegs, lest the iron nails could be drawn from the timber.¹

Such was the state of knowledge mingled with fable when William Gilbert was born in 1544. He made many careful experiments which enabled him to distinguish the true from the false, and writes very scathingly of the 'figments and shrouds, which in the earliest times, no less than now-a-days, used to be put forth by rare smatterers and copyists to be swallowed of men.'

The passages quoted are from the article on 'Magnetism' in the *yc. Britt.*, where they are mentioned as occurring in the translation published by the Gilbert Club) of Gilbert's *De Magnete*.

Gilbert himself was a pattern of scientific honesty, setting down nothing as a fact unless he had verified it by experiment. For instance, he carried out important experiments to show that the earth is a magnet. A full account of his work was given in his book, published in 1600, 'De magnete, magneticisque corporibus, et de magno magnete tellure'—*De magnete*, for short.



DR. WILLIAM GILBERT, PHYSICIAN TO
QUEEN ELIZABETH

*By courtesy of the Director of the Science Museum,
South Kensington*

He was a doctor, and in 1601 was appointed physician to Queen Elizabeth. In our earlier book there is a reproduction of a picture showing her and some of her ladies giving careful attention to a physics lesson.

The present chapter will consist of a review of the more elementary work, though in a few cases this will now be dealt with in greater detail.

Magnetic and non-magnetic substances. If a substance is attracted by a magnet it is said to be *magnetic*; otherwise

is *non-magnetic*. Iron (including steel) is the commonest magnetic substance. Nickel and cobalt, elements belonging to the same chemical family as iron, are also magnetic, but much less so than iron.

In 1903 F. Heusler prepared an alloy of copper, manganese and aluminium which was strongly magnetic. On the other hand, during the war of 1914–18 Sir Robert Hadfield succeeded in making an alloy of steel and manganese which was non-magnetic.¹ Still, it remains broadly true that iron and steel are strongly magnetic, and that other substances are not.

The medium. The medium through which magnetism usually acts is, of course, air. It is easy to show by experiment that it acts readily through glass, wood, most metals—though any common substance in fact *except iron*. Thus

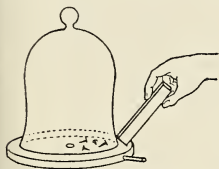


FIG. 31/1

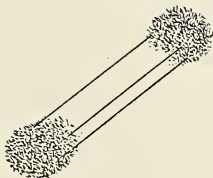


FIG. 31/2

when it is required to screen a substance from magnetic influence it is surrounded by an iron or steel casing.

Magnetism also acts quite readily through a vacuum. This can be shown by putting a few iron tacks under the receiver of an air-pump, near the edge, and then pumping out the air. The tacks are easily made to move about by a magnet held outside (Fig. 31/1).

Magnetic poles. In a magnet as ordinarily magnetised, there are two points, near the ends, where most of the magnetic force seems to be concentrated. This is easily seen by laying a bar magnet, or a horse-shoe magnet, on a sheet of paper on which iron filings have been thickly sprinkled, and then lifting it up. Thick clusters of filings are found to cling near the ends, and very few anywhere else (Fig. 31/2). These points where the magnetic force seems to be concentrated are known

The object was to make it impossible for the Germans, merely by using a compass needle, to find when their vessels were nearing our mines.

as the **poles** of the magnet. The line joining the poles known as the **magnetic axis**.

Direction. Fig. 31/3 shows a magnetised knitting-needle NS mounted so as to be able to swing freely in a horizontal plane. (It is attached at A by a little plasticine to the rounded end of a test-tube, the latter being supported on the sharp point of a knitting-needle AB a little longer than itself.) It is found that the needle always sets in a particular direction, making (in England) a small angle with geographical north and south. We say that it sets 'in the magnetic meridian', a term which will be formally defined a little later (p. 376).

Attraction and Repulsion. In Fig. 31/3 the pole N which points towards the north is known as the *north-seeking pole*, the other one being the south-seeking pole. 'Like' poles (e.g. two north-seeking poles) repel one another, while unlike poles attract.

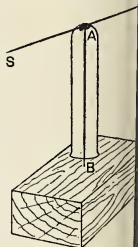
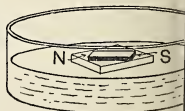


FIG. 31/3

For shortness, a north-seeking pole is usually called simply a *north pole*, but we must be careful to avoid confusion. Since such a pole is always attracted towards the northern part of the earth, and it is unlike poles that attract one another, it follows that the northern part of the earth behaves like the *south* pole of a magnet. When we speak of the earth's magnetic north pole, it is its northerly *position* we have in mind. Compared with magnets in the laboratory, for instance, it is really a south pole.

Poles of equal strength. If we float a block of wood in a trough of water, and across it lay a bar magnet NS, the block turns round until NS is pointing in the usual direction. It does not move bodily however to one side of the trough, and this shows that the poles must be of equal strength. If, for instance, the south pole were stronger than the north, the block would soon drift towards the south side of the trough.



It is interesting to notice that the earliest mariner's compass was very

FIG. 31/4

such as represented in Fig. 31/4. An Arabic writer in a work of date 1282, says that the needle was placed inside a reed, or a splinter of wood, and floated in a basin of water. In rough weather, however, it was probably useless.

Induction. If an iron nail n_1 is made to touch the end of a magnet AB (Fig. 31/5), it sticks to it. There is nothing remarkable about that. If, however, a second nail n_2 is made to touch the lower end of n_1 , n_2 sticks to n_1 . Other nails n_3 , n_4 may be added to the chain—half a dozen perhaps, the number depending chiefly on their weight and on the strength of AB.

If we grip the nail n_1 between thumb and finger and gently remove AB, the lower nails will at once drop off, though perhaps n_2 will continue to cling to n_1 .

Now, in the first instance, if the magnet AB had not been present, n_2 would not have shown the least tendency to cling to n_1 . It is owing to the presence of AB that n_1 has been turned into a magnet. Then, in turn, n_2 has acquired its magnetic proper-

ty from n_1 , and so on through the chain. We say that n_1 has been magnetised *by induction* from AB, n_2 by induction from n_1 , and so on. Magnetism by induction can take place without actual contact (Fig. 31/6), but in that case the effect is feebler, and the chain of nails would be shorter.

Is the induced magnetism temporary or permanent? Our experiment indicates that some of it is temporary, because the lower nails dropped off on removing AB, but that some is more or less permanent, because n_2 remained sticking to n_1 .

We get further light on the question if we repeat the experiment using (i) small pieces of *steel*—pen-nibs or bits of clock-spring will do very well (Fig. 31/7a) and (ii) thin pieces of soft iron of about the same weight. These may be taken from a cocoa-tin (Fig. 31/7b). We find that we can get a much longer chain of soft iron than of steel. Thus the soft iron n_2 in Fig. 31/7b must be a stronger magnet than

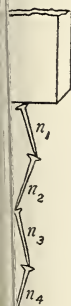


FIG. 31/5

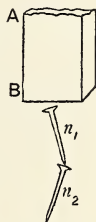


FIG. 31/6

the steel n_2 in Fig. 31/7a, and so *soft iron is more strongly magnetised by induction than is steel.*

But now let us gently remove AB. In the case of the steel nibs, n_2 will still cling to n_1 . In the case of the strip of soft iron, there will be complete collapse. Thus *steel retains its magnetism much better than soft iron.*

The facts just mentioned are often expressed by saying that soft iron possesses greater *permeability* than steel, but less *retentivity*.

Testing for Polarity. If we bring the north pole of a bar magnet near the south pole of a magnetised needle (Fig. 31/8) it will certainly attract it; but if we bring it near the north pole it will not certainly repel it—it may attract it!

The reason is that the north pole of the bar magnet *induces* a south pole in the near end of the needle, and (if the bar is a much stronger magnet than the needle, as it probably is) this induced south pole more than neutralises the original north. Thus we now have unlike poles near to one another, and attraction follows.

Attraction, then, is not a sure proof that two poles were originally unlike. To test our needle for polarity, we must bring up our testing magnet *slowly* to each end of the needle in turn. One end will then be repelled, and that, of course, is the end with polarity similar to that of the near end of the bar magnet. If in spite of bringing up our testing magnet slowly, both ends are attracted, it means that the needle, though made of a magnetic substance, has not been magnetised. Finally, if neither end of the needle is attracted, we should conclude that it was made of some non-magnetic substance such as aluminium.

How to make magnets. Two methods of making a magnet are sufficiently indicated by Figs. 31/9a and 31/9b.

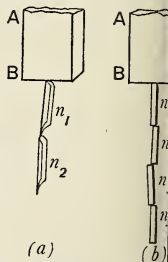


FIG. 31/7

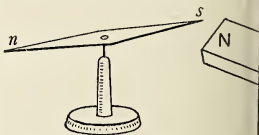


FIG. 31/8

st is known as the 'single-touch' method, the second as at of 'double' or 'divided' touch. In both cases a south pole formed at B. We shall see later that a weak magnet may made by the help of the earth's field.

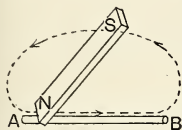


FIG. 31/9 (a)

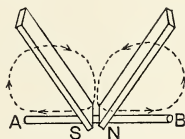


FIG. 31/9 (b)

The modern method of making a magnet, however, is by means of the electric current (Fig. 31/10, where the coil is insulated wire). If you put arrow-heads at the ends of the wire (Fig. 31/11), it will remind you that an *anticlockwise* current is needed to give a north pole. Similar help is obtained from the letter S (clockwise current).



FIG. 31/10



FIG. 31/11

molecular theory of magnetisation. Straighten out a piece of wire-spring about 4 inches long and magnetise it (Fig. 31/9a), necessary testing for polarity. With a pair of pliers break the wire into halves,

test each half for magnetism. It will be found that each half is a magnet. The original poles are not lost, but new poles have appeared at the point of division. (Fig. 31/12b).

Break each of the two magnets into halves. On examination it will be found that there are now four magnets, with polarity as shown in Fig. 31/12c.

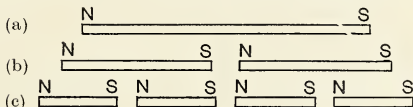


FIG. 31/12

In practice, of course, the bits of spring soon become too short to be broken further, but it seems reasonable to suppose that if we could overcome the mechanical difficulties we could just go on and on making new magnets until we reached very small particles indeed.¹

We can now get a good idea of what happens when we magnetise a piece of steel. Fig. 31/13 represents the unmagnetised steel. Every particle is a magnet, and for convenience the two poles have been distinguished by marking one end with an arrow. There

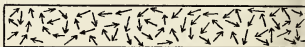


FIG. 31/13

is no orderly arrangement, but on the principle that 'unlike poles attract', most of the particles would no doubt form themselves into closed chains of three, four or more, the north pole of one particle touching the south pole of the next. Such a closed chain would show no free magnetism, because each 'north' would be neutralised by the 'south' that touches it. Again, even if we consider 'free' particles, there would be just about as many pointing in one direction as in another, and so the bar as a whole would show no polarity.

Now suppose we stroke the bar with a magnet as shown in Fig. 31/9a (p. 367). The closed chains are broken up, and the particles set themselves somewhat as shown in Fig. 31/14, their south poles all pointing

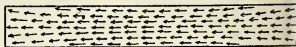


FIG. 31/14

towards the 'north,' which is passing over them. At one end the north poles are all pointing in the same direction, with their south poles to neutralise them, and so at that end the bar shows north polarity. In a similar way it shows south polarity at the other end. The central parts of the magnet show no polarity, because the north pole of one particle is touching the south pole of the next.

¹ The very smallest particles would be molecules, but there is some reason to believe that the limit would be not the molecule itself, but a cluster of molecules—what we have called a 'very small particle' indeed.'

If we break our magnet into two, new poles are found at each broken end because the particles were already pointing in one direction—and so on, no matter how many times we break it.

Confirmation of the theory. If it is true that a magnet owes its properties to a special arrangement of its particles, we ought to find that if it is knocked about violently it should lose its magnetism, because the knocking would upset the orderly arrangement.

That is just what we do find. The more a magnet is knocked about, dropped on the floor, etc., the weaker it becomes.

Again, *heating* would produce violent agitation of the particles, and so by shaking them out of their orderly arrangement should produce demagnetisation.

That also is found to be the case. In fact, if a magnet is made red hot and then cooled, it will be found to have lost its magnetism completely.

Our theory suggests that there is a definite limit to the force to which a piece of steel can be magnetised, the limit being reached when the magnet particles have all attained the end-on position. This also is borne out by experiment. In magnetising a piece of steel, a point is soon reached at which it is 'saturated.'

Keepers. It is found that a horse-shoe magnet retains its magnetism much longer if when it is put away, a short bar of soft iron called a 'keeper' is placed across its ends (Fig. 15a). We can now see the reason. For clearness, the diagram has been drawn on a large scale (Fig. 31/15b). North poles are induced in the ends of the keeper, and the induced north poles at the ends of the keeper are now held in position by the magnet-particles at the end of the keeper; similarly, of course, at the end S.

To apply the 'keeper' principle to bar magnets, they are always kept in pairs with their poles in opposite directions (Fig. 31/16) and keepers are placed across the ends.

Demagnetising. We have seen that a magnet may be demagnetised by heating it or by knocking it about violently, the effect in either case being to disturb the orderly arrangement of the magnet-particles. A much better method,

however, is to use the apparatus already described for magnetising (Fig. 31/10), except that now we *must* use an alternating current. The magnet is placed inside the coil (Fig. 31/17) the current is turned on, and then the magnet is slowly

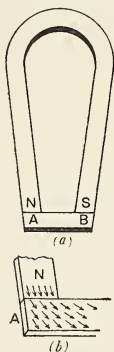


FIG. 31/15



FIG. 31/16

withdrawn. As the current surges first one way and then the other, the polarity of the magnet is constantly reversed. At A, for instance, we have first a north pole and immediately afterwards a south pole. But as the magnet is withdrawn it is less and less exposed to the magnetising influence of the current, and so it becomes weaker and weaker. When it has been taken completely out of the coil and some distance beyond, the influence of the current becomes practically nil and the demagnetisation is complete. It is a good plan, however, to repeat the process with the magnet in the reversed position (i.e. A to the right).

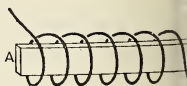


FIG. 31/17

Fields of Force. The influence of a magnet is, of course, noticeable for some distance round it—iron filings are attracted, or a compass needle is made to alter its direction. In the quaint language of Gilbert, 'From about a magnet

dy, the virtue magnetical is poured out on every side round in an orbe.' The space throughout which the influence of a magnet can be detected is known as its *field of force*.

A complete knowledge of the field of force of a magnet would mean that at any point in the field we know (i) the *direction* of the force and (ii) its *magnitude*. It is much easier to obtain information about the first than about the second, and it is direction with which we shall be mostly concerned in the present chapter.

Lay a strong bar magnet on the table, and over it lay a sheet of drawing-paper. It is better to keep the paper level, and this

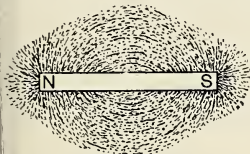


FIG. 31/18

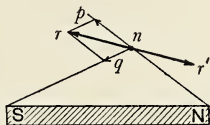


FIG. 31/19

be done by placing rulers, three-ply or something of the sort over its edges. Sprinkle iron filings over the paper as evenly as possible, using a pepper-pot. Now with the end of a pencil, tap the paper gently in different parts. The filings quickly set themselves in curves (Fig. 31/18).

Let us try to make out what has happened. Each filing, because it is in the neighbourhood of a magnet, becomes a magnet itself by induction. Suppose n is the north pole of one of these small magnetised filings (Fig. 31/19). It is pulled from N by a force which we will represent by the line np . It is attracted to S by a smaller force represented by the shorter line nq (the force being smaller because the distance is greater). We complete the parallelogram of which np and nq are two adjacent sides and draw the diagonal nr . nr represents the resultant force on the north pole of the magnetised filing. The resultant force on its south pole p can be represented by nr' , equal and opposite. If free to move, the filing will evidently set itself in the direction rnr' . To give it freedom of movement that we tapped the paper. This caused the filing to give a little jump into the

air, and set it free for a moment from the frictional resistance of the paper.

For *every* filing, then, there is a definite direction in which it will arrange itself if free to do so; and when each filing has so arranged itself, the result is the curves we have obtained in our experiment.

Going back to Fig. 31/19, we can see that if n were a free north pole, it would begin to *move* in the direction nr . A moment later, because it would be in a slightly different position, we should obtain a slightly different parallelogram, and its direction of motion would be altered a little. In fact it would move along the curve on which it lies. Such a curve is known as a **line of force**, which we may formally define as *the line along which a free north pole would travel as the result of magnetic forces acting upon it*.

Actually we cannot obtain a free north pole—there is always a south at the other end of the magnet. However, we can so arrange matters that the influence of the south pole is scarcely felt.

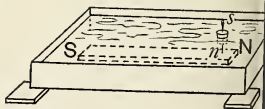


FIG. 31/20

A sewing-needle is strongly magnetised and passed vertically through a small cork, which is floated on water in a shallow trough. One end of the needle, say the north end, should nearly touch the bottom. The trough is supported just above a strong bar magnet, the north pole of which is the same as that of the needle. It will be found that the cork moves along a curved path similar to those we have been considering.

You should now use the iron filings method to obtain records of the various cases represented by Figs. 31/20 to 31/26, afterwards comparing your results with the illustrations. (N.B. In these experiments, Robinson's *ball-ended magnets* are sometimes used instead of bar magnets. They have the advantage that as the magnetic pole is at the centre of the sphere, there is no vagueness about its position.)

Neutral points. Notice the point marked X in Fig. 31/24. If we imagine a free north pole in that position, it would not move, because it is equally repelled both from right and left. It is known as a **neutral point**, which we may define as *a point subject to magnetic forces of which the resultant is zero*.

Other neutral points are indicated in Figs. 31/25 and 31/26. The presence of a piece of soft iron in a magnetic field produces very marked changes. Figs. 31/27 and 31/28



FIG. 31/21

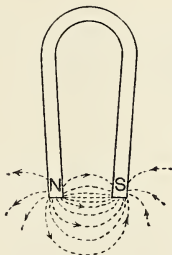


FIG. 31/22

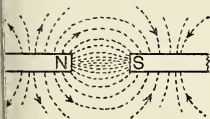


FIG. 31/23

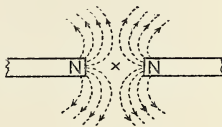


FIG. 31/24

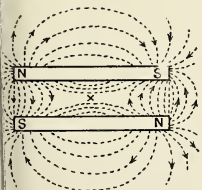


FIG. 31/25

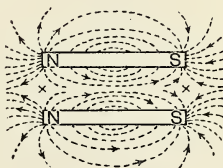


FIG. 31/26

ould be compared with Fig. 31/23. The lines of force are ner actually drawn into the iron, or are at least bent towards

Inside the ring (Fig. 31/28) we have no lines of force. Another way of considering these cases is to remember

that the soft iron becomes a magnet by induction (the poles are marked with small letters). Each case then reduces to that illustrated in Fig. 31/23, except that the proximity

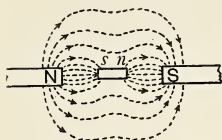


FIG. 31/27

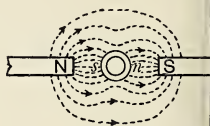


FIG. 31/28

unlike poles occurs twice over in Fig. 31/27 and also in Fig. 31/28.

In Fig. 31/28 we should say that the ring serves as a *magnetic screen*. We have a very good example of magnetic screening

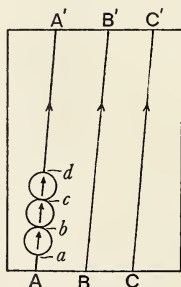


FIG. 31/29

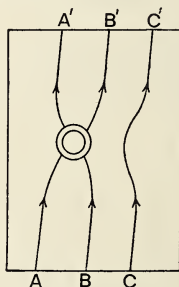


FIG. 31/30

in the case of a submarine. The lines of force due to earth cannot penetrate the steel shell of the vessel, and so the needle of an ordinary compass would not point in any particular direction. A gyro-compass, which works on an entirely different principle, is used instead.

Earth's field. The earth's field is too weak to cause iron filings to arrange themselves along lines of force. However, a small compass needle will give us the direction of the lines of force at any particular point such as A (Fig. 31/29),

moving the compass steadily forward from ab to bc , bc cd and so on, we can obtain a line of force of any desired length. We can then obtain other lines, BB' , CC' . . . by starting from B , C . . .

It is interesting to repeat the experiment, but this time putting a solid or hollow iron cylinder on its base in the middle of the paper. We notice how the lines tend to be drawn into the cylinder (Fig. 31/30).

Magnets in Earth's Field. Starting with a fresh piece of paper, a strong bar magnet in the middle of it, with its south pole pointing (magnetic) north. Use your compass needle to obtain a diagram of the field, and compare your results with Fig. 31/31.

Notice the neutral points X_1 and X_2 . Imagine a free north-seeking pole at X_1 . On balance, it is attracted by the bar magnet, because the repelling force of N , which is distant, is not so strong as the attractive force of S , which is near. But its action to the bar magnet is balanced by the attraction it experiences in the opposite direction, from the earth. Hence the resultant of the magnetic forces is zero. X_1 is a neutral point. The case of X_2 is similar. A free north pole placed there would be, on balance, repelled by the magnet but equally attracted by the earth.

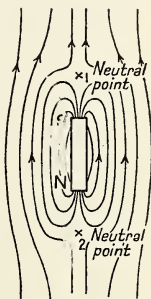


FIG. 31/31

X_1 and X_2 are equidistant from the earth's north magnetic pole (a difference of two feet or so, in a matter of some hundreds of thousands of miles, 'doesn't count'). Hence a free north pole at X_1 would experience the same attractive force from the earth as at X_2 . Therefore, to balance, it must experience an equal force in the opposite direction from the magnet—and X_1 and X_2 must be equidistant from the respective ends of the magnet.

Declination. Most people know how to find the Pole Star. When looking at it you are looking almost exactly north. However, you possess a pocket compass, you will find that the needle does *not* point to the north, but 10° or so to the left (west) of it. We say that 'the *declination* is 10° '

West.' This term is related to two others (*geographical meridian* and *magnetic meridian*) and it will be convenient to define all three of them now.

The geographical meridian at any point is the plane which passes through that point and the earth's geographical north and south poles.

The magnetic meridian at any point is the vertical plane passing through the axis of a compass needle placed at that point.

The angle between the geographical meridian and the magnetic meridian is known as the declination (or variation).

These points are illustrated in Fig. 31/32. For simple practical purposes we may think of the declination as being the angle which our compass needle makes with a line running in a geographical north and south direction.

Sailors and airmen, and often travellers on land, make use of a knowledge of declination (or variation, as it is still sometimes called). The compass gives them *magnetic* north, and from this, knowing the declination, they obtain *geographical* north—the north of the map. At one time it was thought that declination was the same at all points on the earth's surface. Columbus in 1492 was possibly the first to discover that this was not the case, and no doubt this added to the troubles of a very worrying voyage. He was certainly the first to find a place where the declination was zero (not far from the island of Corvo, one of the Azores).

Maps can be obtained in which lines are drawn through points which have the same declination. These are known as *isogonic lines*. The special isogonic line joining up places where the declination is zero is known as an *agonic line* (actually there are two of these lines).

Declination not only varies from one place to another, at any one place it changes with time. In London, for instance, it was $11^{\circ} 15'$ E. in 1580, zero in 1657 and 24° W. in 1818, since when it has been steadily decreasing. In fact, a captain using an old set of tables might easily run his ship on the rocks.

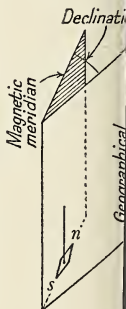


FIG. 31/32

Inclination or Dip. The earth behaves in many ways as if there were an immense bar magnet inside it, passing through the centre and making an angle with the line joining the geographical north and south poles (NP and SP in Fig. 31/33). This is, of course, not the case—for the thing, the earth's interior is far too hot—but the idea is often helpful. Such a magnet would have lines of force somewhat as shown in the figure. At some places, such as A, they would be parallel to the earth's surface, while at others, B and B', they would be at right angles to it.

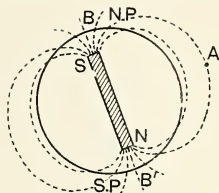


FIG. 31/33

At other points they would be inclined at all sorts of angles between the extremes of 0° and 90° . In London, for instance, the lines of force actually make an angle of about 68° with the horizontal. If we can imagine a compass needle with a sort of universal joint for its pivot, it would (in London) not only point magnetic north, but it would point downwards at an angle of 68° . Our imaginary 'universal joint' needle would still be swinging in the plane of magnetic meridian, but it would be inclined to the horizontal at the same angle as the earth's lines of force.

Now if a needle is to be free to align itself to the horizontal, it could swing in a vertical plane, and this is the essential feature of the dip needle as it is called (see Fig. 31/34).

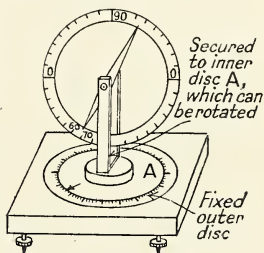


FIG. 31/34

The *inclination* or *angle of dip* is the angle made with the horizontal by a magnetised needle, pivoted at its centre of gravity and swinging freely in the magnetic meridian.

To find this angle we have to set the instrument so that the graduated circle, over which the needle moves, lies in the plane of the magnetic meridian. We might obtain this position by using an ordinary compass needle, and then by

'sighting' arrange that our vertical circle in the edge-position just covers the line of the needle. This would involve using a second instrument, but actually we can obtain more accurate results without doing so.

In Fig. 31/35 suppose M and R are adjacent sides of an ordinary box. Clearly the two planes are at right angles. Suppose M is in the plane of the magnetic meridian, and then AT represents the earth's magnetic force in that plane, both in magnitude and direction (we shall discuss the question of magnitude in the next chapter). If from A we draw a horizontal line AH and a vertical line AV, and complete the rectangle AHTV, we may regard AT as the resultant of a horizontal component AH and a vertical component AV.

Now suppose our dip circle is placed so as to lie in the plane R, at right angles to the plane M. N is the north-seeking pole of the needle, S the south-seeking pole. What will happen?

The horizontal component of the earth's magnetic force will attract N, and repel S, in the directions indicated by the dotted lines. But the needle is so mounted that N and S can make no movement in these directions. Thus the horizontal component produces no effect except to cause a slight strain on the bearings.

The only effective force is the vertical component, which will pull N vertically downwards. Evidently *when the plane of the dip circle is at right angles to the magnetic meridian, the needle will set vertically*. In any other position the horizontal component would have some effect, and the needle would not set vertically. It is assumed, of course, that the needle is pivoted at its centre of gravity; otherwise it would tend to set vertically, quite apart from magnetic forces.

In using the instrument, then, we first see that the box is level (a good instrument includes a spirit level). We then rotate the base of the vertical pillar until the scale reading is

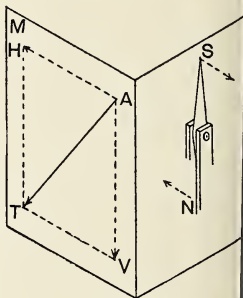


FIG. 31/35

is that the needle is vertical. As just explained, the plane of the dip circle is now *at right angles* to the magnetic meridian, and so by turning the base another 90° , we get it into the required plane. We then take our reading, both ends of the needle, to eliminate any error caused by the needle not being pivoted exactly at the centre of the scale.

We then turn the vertical pillar through an angle of 180° and repeat the two readings. This is to eliminate any error caused by the zero line of the scale not being horizontal. We now take the mean of our four readings.

To illustrate dip, we may suspend an unmagnetised knitting-needle with fine cotton, so that it hangs horizontally. Its centre of gravity will, of course, be in the line of the cotton. With two bar magnets, using the method of double touch (Fig. 31/9*b* on p. 367), we then magnetise the needle, taking care not to let the cotton slip along the needle. The latter will now be found to tilt with its north pole downwards.

An early experiment. One of the first men to study the question of magnetic dip was a compass maker named Robert Norman who lived in Wapping. He took great care to balance his compass needles as exactly as possible at their centre of gravity, so that they should be level. When he afterwards magnetised them, however, the north end always tilted slightly downwards. He soon suspected the cause, and actually constructed a dipping-needle with which he found the angle of dip to be $71^\circ 50'$. This was in 1576.

Five years later he wrote a book called 'The Newe Attractione.' He argued that somewhere under the earth there must be a 'poynt Respective' towards which the needle turned, and by finding the direction of the needle at various parts of the earth's surface, and noting where the lines intersected, he should be able to find it.

Norman, of course, was not in a position to do this, because the necessary information about the angle of dip was not available. It is quite easy, however, to do it now. The angles of dip for some different latitudes on the meridian of Greenwich are given at the top of p. 380.

Fig. 31/36 will suggest how to go to work. N and S represent the (geographical) north and south poles, and CQ a part of the equator (lat. 0°). The latitudes are measured

	North								South				
Latitude (°)	65	60	50	40	30	20	10	4	0	20	40	60	7
Inclination (°)	75	72	68	59	48	31	10	0	10	38	54	62	9

as *angles* from CQ (only a few shown). Now to represent, for instance, the information 'Lat. 60° , Inclination 72° .' Join C to the '60' point, K. This line is the vertical through K, i.e. it points to the earth's centre. Draw the tangent KH. TH is the horizontal at K, and the dipping needle would point as shown.

Do your arrows give any rough idea of the position of the earth's magnetic poles?

Isoclinic Lines. We have seen that the inclination at London is 68° . There are many other points on the earth's surface which have the same inclination—in Germany, Russia, Siberia, the United States and elsewhere. If we draw a line joining up all points having the same inclination we obtain what is called an *isoclinic line*, or *isoclinical*.

An *acclinic line* is a line joining points where the angle of dip is zero.

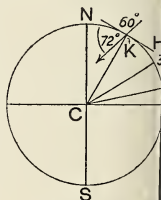


FIG. 31/36

On the theory that the magnetic properties of the earth are such as would be produced by a magnet placed somewhat as in Fig. 31/33 we should expect to find magnetic north and south poles—points where the dip would be 90° —in the far north and far south respectively, with an acclinic line following roughly the direction of the earth's equator. Isoclinic lines we should expect to follow, roughly, lines of latitude.

Actually a south magnetic pole has been located in lat. $71^{\circ} 10' S.$, long. $150^{\circ} 45' E.$ (Mawson, 1914) and a north magnetic pole in lat. $73^{\circ} 15' N.$, long. $94^{\circ} 30' W.$ ¹ Comparison

¹ The latter position is given by the Canadian Dominion Observatory. It is deduced from observations made during the flight of the A-1 (R.A.F., under Wing Commander McKinley, May 1945), together with land observations made later. For this information the author is indebted to the courtesy of the secretary of the Royal Geographical Society.

with earlier observations suggests that these positions alter rather rapidly. Thus Roald Amundsen in 1903-06 gave the position of the north magnetic pole as 71° N., 96° W., while in 1831 it was $70^{\circ} 05'$ N., $96^{\circ} 46'$ W. (James Clark Ross).

The aclinic line does indeed follow roughly the direction of the equator, never being more than about 10 degrees north or south of it. The isoclinic lines cannot be said to follow very closely the lines of latitude, especially in the southern hemisphere. However, their general direction is east and west.

No doubt the 'single bar magnet' theory is too crude to lead to very accurate results. By supposing a second magnet to be present, we find that theoretical and actual values for the angle of dip at different parts of the earth's surface agree much better, and the agreement can be still further improved by increasing the number of imaginary magnets. At the theory then becomes too complicated to be very useful.

Near the Magnetic Pole. The behaviour of a mariner's compass in the neighbourhood of the earth's north or south magnetic pole might be expected to be troublesome. One needle would have a strong tendency to turn downwards. Further, the tendency of a compass to point north and south depends on the *horizontal* component of the earth's magnetic force (cf. Fig. 31/35 on p. 378), and near the poles this component is nearly zero.

Keeping these points in mind, it is interesting to note the following passage from a letter written in 1911 by Lieut. R. Bowers.¹ 'The compass required standing still to look at it every time. Our sledging compasses are spirit ones, and as steady as a small hand compass could possibly be. You will understand, however, that owing to the proximity of the magnetic Pole the pull on the needle is chiefly downwards. It is forced into a horizontal position by a balancing weight on the N. side, so it is obvious that its direction power is greatly reduced. On the ship, owing to vibration of the engines and the motors, we were absolutely unable to steer

¹ In company with Capt. Scott, Wilson, Oates and Evans he reached South Pole on 18th January, 1912, but the whole party perished on the return journey.

by the compass at all when off the region of the Magnetic Pole'.¹

We have a similar effect, of course, in the neighbourhood of the north magnetic pole. Cf. the following little passage from 'Plowing the Arctic' by G. J. Tranter '... the weather had suddenly become dirty with a strong wind, and it was unsafe to proceed, for now the compass was entirely unreliable owing to the proximity of the Magnetic Pole, and they had to be able to see their way, needing the land for a guide.'

Magnetising Effect of Earth's Field. We know that a piece of iron or steel, in the neighbourhood of a magnet, becomes a magnet itself by induction (p. 365). We should therefore expect that an iron rod which has been lying in, or nearly in, the magnetic meridian would be magnetised, and especially if it has been sloping at an angle of about 68° .

Actually we are much more likely to meet with iron rods (railings, for instance) 'sloping' at 90° than at 68° . It is interesting, using a pocket compass, to test these for polarity (the compass must, of course, be brought up *slowly*). The lower end will usually be found to repel the north-seeking end, showing that it is north-seeking itself. Garden spades, forks, etc., may also be examined. The writer found that his garden spade, and every prong of his garden fork, showed north polarity at the bottom, as did two lengths of 'arm iron,' which had been placed in an upright position for fence

Questions

1. Make a list of the chief properties possessed by a magnet and describe how you would demonstrate each of these properties.

How would you show that the poles of a magnet are of equal strength but of opposite polarity? *C.W.B.*

2. A magnet usually attracts a piece of soft iron; repulsion is a property observed with permanent magnets. Explain. *O.F.*

3. How would you magnetise a bar of soft iron by (a) suitable stroking the bar, (b) a method of magnetic induction, (c) using an electric current? Illustrate by diagrams which indicate the polarity acquired.

What do you consider takes place inside the iron bar during magnetisation, and how do you explain the fact that there is a limit to the intensity of magnetisation of the bar? *Dur.*

¹ Quoted from *The Worst Journey in the World*, by A. C. G. Garrard.

. Explain the following:—

- 7) If the N. pole of a strong magnet is held at some distance from the N. pole of a weak magnet, it repels it, but if it is brought close to it, it attracts it.
- 8) It is possible to magnetise a steel knitting needle so that there are north poles at both ends, and a steel ring so that it has no poles at all.
- 9) When a magnet is broken or cut into two pieces, each piece is a complete magnet. *Camb.**

How would you demagnetise a bar magnet?

. Explain the terms *pole of a magnet*, *magnetic line of force*, *magnetic field*, *neutral point*.

Describe how you would plot the lines of force in the neighbourhood of a bar magnet, placed with its axis in the magnetic meridian, and with its North pole pointing south.

Draw a diagram showing the lines of force, marking their directions and indicating the positions of the neutral points.

C.W.B.

. Explain the meaning of the terms *magnetic variation* (or *magnetic declination*) and *magnetic inclination* (or *dip*).

Describe how you would measure *one* of these quantities in the laboratory.

How can an approximate magnetic model of the earth be made by suitably placing a magnet inside a sphere. Give a diagram. *J.M.B.*

Explain why the needle of a dip-circle points vertically upwards when the plane of the dip-circle is at right angles to the magnetic meridian.

'The latter [i.e. a knitting-needle suspended at its c.g. and magnetised] will now be found to tilt with its north pole upwards' (p. 379).

What difference, if any, would you expect to observe if you carried out this experiment at some place in the southern hemisphere, e.g. Melbourne?

If you were provided with an unmagnetised steel knitting needle, how would you use it to demonstrate magnetic dip? Give details of your method of magnetising the needle.

Explain, giving reasons, how you would magnetise a piece of steel as strongly as possible in the earth's magnetic field. *C.W.B.**

CHAPTER 32

MAGNETIC MEASUREMENTS

IN his age-long history man has always been engaged exploring the world around him. He has been interested in finding where rivers have their source; he has wanted to know what lay on the other side of a range of mountains or beyond the ocean. In his ceaseless enquiries he has succeeded so well that, in its main outlines, geographical exploration may be regarded as complete.

But geographical exploration is only one part of the quest. We see the point very well illustrated in Scott's expedition to the Antarctic, already referred to in these pages. Certainly his work had a geographical side, but he took with him a remarkable team of highly trained scientists, who brought back with them a wealth of information about wind velocity and air temperatures, atmospheric humidity, the depth and salinity of the ocean and the plant and animal life contained in it. The magnetic laboratory was set up in a cave in a small glacier. There, by an ingenious arrangement which included a clockwork mechanism, they obtained continuous records of the direction and intensity of the earth's magnetic force. There, too, they made careful measurements of declination and inclination, and of other values which we shall be considering in the present chapter. The illustration (from *The Great White South*, by H. G. Ponting) shows Dr. Simpson making observations with his magnetometer.

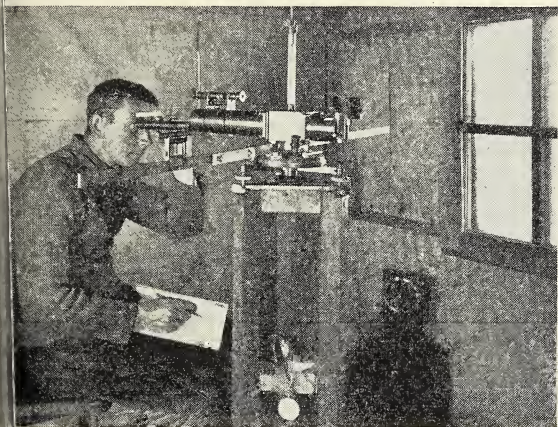
Let us get rid of the idea—if we have it—that the study of magnetic measurements is 'just school stuff.' It is school stuff, of course, but it is exploration stuff as well, and the more we know of it, the deeper will be our interest in the story of world exploration.

Pole strength. As is always the case, we find that we cannot make measurements until we have fixed upon a suitable unit, and we shall begin by considering a unit in which we can express the strength of a magnetic pole.

We know that like poles repel one another with a certain force. The stronger they are, and the nearer they are together,

stronger the force of repulsion. Now, we already have a unit of force, the *dyne* (p. 33), and this helps us to fix the new unit we are considering. *Unit magnetic pole* is one of such strength that it will repel a similar pole, placed one centimetre away from it in air, with a force of one dyne. This unit pole has been given the name *Weber*.¹

We shall have need almost at once of another definition



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DR. SIMPSON MAKING MAGNETOMETER OBSERVATIONS
IN THE ANTARCTIC

connected with unit pole. The *intensity of a magnetic field at any point* is measured by the force which would be experienced by a unit north pole placed at that point. The name given to this unit is the *oersted*¹ or *gauss*.¹ In London, for instance, a unit north pole would be attracted with a force of about one dyne. We say, then, that the intensity of the earth's magnetic field, in London, is 0.18 oersted.

It is implied in the definition of the *weber*—even if we had not known it already—that the force of repulsion varies with

Weber, *Oersted* and *Gauss* were eminent 19th-century physicists. The first was a Dane; the other two were Germans.

the distance. But to what extent? If the distance doubled, is the force simply halved, or what? That is a question we must try to answer. In doing so, we shall make use of an instrument called a *magnetometer*, or *deflection magnetometer*, and we will consider this instrument first.

Deflection Magnetometer. Suppose a short magnetic needle has a bar magnet placed on its east or west side. In Fig. 32/1 this is supposed to be well to the east (lying on an east-west line passing through the centre of the needle), but too far off to appear in the diagram. Its south pole faces the needle. Because of its distance, and the shortness of ns , we may suppose it to be opposite either n or s .

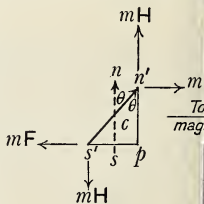


FIG. 32/1

Let H be the field due to the earth. Then unit pole would experience an attractive force H , so if m is the strength of the pole at n' , the attractive force is mH . At s' we have an equal and opposite force. These two forces make up a couple (p. 82), with a moment or turning power of $mH \times ps'$.

Similarly the magnet (on the east) gives rise to a couple of moment $mF \times pn'$, where F is the field at (and very near) n' due to the magnet.

As the needle is stationary, the two moments must balance, so $mH \times ps' = mF \times pn'$, or $F = H \times ps'/pn' = H \tan \theta$.

We will now notice how the magnetometer is constructed, and we shall then see that this relation $F = H \tan \theta$ will enable us to answer our original question—viz., how is the force between two magnetic poles related to the distance between them?

The central feature of the magnetometer is a *small* magnetic needle ab pivoted on a vertical support (Fig. 32/2). It must be small so that the field in which it moves may be regarded as uniform; otherwise we should have hopeless complications. But it is much easier to take angular readings with a long needle than a short one, and so a long aluminium pointer crosses the magnetic needle at right angles. Its ends

er a scale showing degrees. Inside this is a circular sheet mirror glass, and when taking a reading we take care that the pointer just 'covers' its image. In this way we avoid the error of parallax. The needle and scale are enclosed in a square frame, opposite sides of which contain slots S_1 and S_2 . Into these slots are fitted two rather long wooden arms so graduated as to show the distance in centimetres from the centre of the instrument. From the end-on view it will be seen how the arm is so shaped that a magnet can conveniently be placed on it, touching the graduation marks. The arrangement is such that the magnet is exactly opposite the centre of the needle.

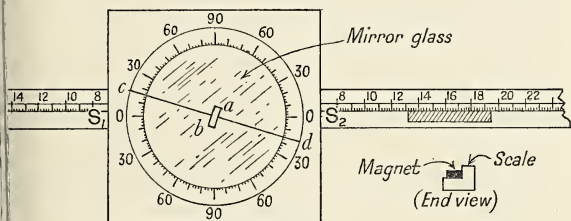


FIG. 32/2

taking care that no magnets or iron work are near, we slowly turn the magnetometer until the aluminium pointer is over the zero of the scale. (There is sometimes a slight difference between the readings of the two ends, due, e.g., to the pointer being bent. In that case the mean of the two readings must be zero.)

Now comes our first real difficulty. We are investigating the field produced at the centre of the instrument by a pole of our magnet. But the magnet has *two* poles, and the second pole will simply upset our results. If we use a really long magnet, however, this second pole will be so far away as not to cause very serious disturbance; so we use a magnet of a thin-needle shape, at least 45 cm. long. Its poles are placed at the ends. We start with one end about 15 cm. from the needle and observe the deflection (reading both ends of the needle and taking the mean). We then repeat at

distances increasing by 2 or 3 cm. each time. Suppose we obtain the following results:—

Distance in cm. (d)	θ	$H \tan \theta$	$H \tan \theta \times d$	$H \tan \theta \times$
15	36.8	0.7481 H	11.2 H	168 H
18	27.1	0.5117 H	9.22 H	166 H
20	22.5	0.4142 H	8.28 H	166 H
22	19.2	0.3482 H	7.66 H	169 H
25	15	0.1679 H	6.70 H	167 H

$H \tan \theta$ is the field at the centre of the circle, due to the magnet ($F = H \tan \theta$); it is the force that would act on a unit pole placed there. When we multiply this force by the distance ($H \tan \theta \times d$), we obtain very variable products (11.2 H, 8.28 H, etc.); but when we multiply it by distance *squared* ($H \tan \theta \times d^2$), we notice that the answers are practically constant, showing that the force varies inversely as the *square* of the distance. This result, known as the **Law of Inverse Squares**, is so important that we had better have a formal statement of it.—*The force of attraction or repulsion between two magnetic poles varies inversely as the square of the distance between them.*

In the experiment just described we get rid of the action of the unwanted pole by using a very long magnet. Another method sometimes used is to arrange a magnet with its unwanted pole vertically above the centre of the magnetometer needle. In that position it is unable to deflect the needle, and so all deflection effects are due to the other pole. Diagrammatically, two positions of the magnet are shown in Fig. 32/3.

The following experiment gives an indirect but very good proof of the law of inverse squares. We begin by *assuming* the truth of the law. We then work out certain results that should follow if the law is true. We prove by experiment that these results *do* follow, and we conclude that our original assumption must have been correct. Let us now consider the experimental work in detail.

Draw the outline of a bar magnet on a large sheet of paper and mark the positions of the poles, N and S. We shall

far out if we put them at one-twelfth the length of the magnet from each end (Fig. 32/4).

Mark any point P on the paper, and *assuming the truth of the law of inverse squares*, find which way a free north pole would move if placed at P. To do so join NP and SP and produce NP to Q. Measure NP and SP and square the results. Suppose $NP = 15.2$ cm. and $SP = 9.8$ cm. We have $2^2 = 231$ and $9.8^2 = 96$.

Now, a free north pole would be repelled along PQ, but attracted along PS. If the forces of attraction and repulsion be *directly* proportional to the squares of the distances, we could mark off a length of 231 units along PQ and 96 units along PS. But the forces are *inversely* proportional, so we

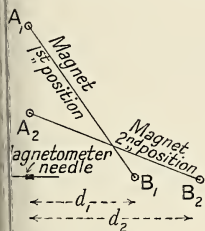


FIG. 32/3

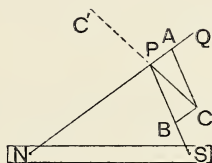


FIG. 32/4

mark off the 96 units along PQ and 231 units along PS. A convenient unit would be, say, $1/40$ th cm. Then 96 units = 2.4 cm. and 231 units = 5.78 cm. In the figure, then, we mark off $PA = 2.4$ cm. and $PB = 5.78$ cm.

PA and PB represent, in magnitude and direction, the forces acting on our free north pole, its *actual* direction of movement would be along PC, the diagonal of the completed parallelogram.

Now, suppose a compass needle to be placed at P. Its north pole will be driven along PC. Clearly the forces acting on its south pole, though the same in magnitude, are opposite in direction, so this south pole would be driven along PC'. The needle should evidently set in the direction C'PC.

Place the bar magnet and place a compass needle at P. You will find that it actually sets along the line CPC'.

This *might* be the result of an odd coincidence, so report the work for several other positions of P. We find that two directions agree every time. Other members of the class all report the same agreement. We are therefore bound to conclude that the assumption underlying all our calculations is true—and that assumption is the Law of Inverse Squares.

Mathematical Expression. In calculations connected with magnetism, we shall find that we are again and again making use of the following statement: *If a magnetic pole of strength m units is d centimetres from another pole of strength m' units, the force of attraction or repulsion between the two poles is $\frac{mm'}{d^2}$ dynes.* (It can easily be shown that this follows from the law of Inverse Squares.)

N.B. Unlike poles are expressed by opposite signs. Thus a north pole of 6 units and a south pole of 8 units would be expressed as $+6$ and -8 . If they were 4 cm. apart, the force between them would be $\frac{(+6) \times (-8)}{4^2} = -3$ dynes.

the minus answer indicating that the force is one of attraction.

Some applications of the Law. We have just had one very simple application—finding the force of attraction between two poles at a given distance apart. Here is a slightly harder one:—

Example 1. *A bar magnet 12 cm. long and of pole strength 36 units is placed horizontally in an east-west position, with its north pole pointing east. Calculate the field strength at a point P, 12 cm. east of the south pole, on the prolongation of the axis.*

The *field strength* at P means the force which would be experienced by a unit north pole placed there (p. 385).

In calculations where there is no statement to the contrary it is to be assumed that the poles are at the ends of the magnet.

Distance of P from S = 12 cm. (Fig. 32/5).

\therefore unit north pole at P would experience a force of $-\frac{36}{12^2}$ dyne.

Distance of P from N = $12 + 12 = 24$ cm.

\therefore unit north pole at P would experience a force of $\frac{36 \times 1}{24^2} = \frac{1}{16}$ dyne.



FIG. 32/5

Both forces act along the line NSP, and the resultant is force of $-\frac{1}{4} + \frac{1}{16} = -\frac{3}{16}$ dyne, the sign indicating that the resultant force is one of attraction. Thus the *strength* of the field at P is $\frac{3}{16}$ oersted. Its *direction* is along PS.

Magnetic Moment. The field produced by a magnet depends partly, of course, on the strength of the magnetic poles. But it also depends very much on the length of the magnet, especially at points in the direction of the axis.

For instance, suppose N_1S_1 is a short magnet and N_2S_2 is a long one. Their poles are of equal strength. P_1 is a point in S_1N_1 produced, and P_2 another point in S_2N_2 produced. P_1N_1 is equal to P_2N_2 .

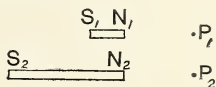


FIG. 32/6

Now, if we had to consider only the poles at N_1 and N_2 , the field at P_1 would be of the same strength as at P_2 , because the distances are the same, and the poles are of the same strength.

But in both cases the strength of the field due to the north poles is *weakened* by the fields due to the south poles at S_1 and S_2 . This weakening effect is not so great in the case of the longer magnet, because S_2 is farther away from P_2 than S_1 is from P_1 . Thus the resultant field due to the long magnet is stronger than that due to the short one.

Because the field produced by a magnet may be affected by its *length* as well as by its *pole strength*, we often have used a quantity which includes both of these. This quantity is known as the **moment** of the magnet, and is obtained by multiplying the pole strength by the magnetic length of the magnet (magnetic length = distance between poles). It can easily be shown that the magnetic moment as just defined is equal to the moment of the couple (p. 82) required to keep the magnet at right angles to a field of unit strength.

Comparing magnetic moments. We have seen that if a short magnetic needle is deflected from the meridian through an angle θ by the action of a magnet placed on its east or west side, then $F = H \tan \theta$, where F is the field produced by the deflecting magnet at the centre of the short needle (386, Fig. 32/1).

Now, if M is the moment of the magnet, and if its length is short compared with d , the distance of its centre from the needle, it can be proved that $F = \frac{2M}{d^3}$. But $F = H \tan \theta$.

Hence under these conditions $F = \frac{2M}{d^3} = H \tan \theta$.

To compare the moments of our two magnets we proceed as follows:—

We set our magnetometer with the pointer at the zero of the scale—i.e. arms pointing east and west, as in Fig. 32 on p. 387. We then place our first magnet so that its middle point is 30 or 40 cm. from the centre, say on the east side of the needle, and observe the deflection θ_1 .¹

We repeat with the second magnet, its middle point being in the same position as before. Call the deflection this time θ_2 .

If M_1 is the moment of the first magnet and M_2 of the second, then in the first case $\frac{2M_1}{d^3} = H \tan \theta_1$, and in the second $\frac{2M_2}{d^3} = H \tan \theta_2$.

$$\therefore \frac{M_1}{M_2} = \frac{\tan \theta_1}{\tan \theta_2}.$$

e.g. if $\theta_1 = 32.3^\circ$ and $\theta_2 = 44.7^\circ$

we should have $\frac{M_1}{M_2} = \frac{\tan 32.3^\circ}{\tan 44.7^\circ} = \frac{0.6322}{0.9896} = \frac{1}{1.57}$.

In the experiment just described we have only compared the moments of two magnets. If the value of H is known we can find the actual value, for $\frac{2M}{d^3} = H \tan \theta$, and if we know the values of H , d and θ we can obviously calculate M . Lastly, $M = (\text{magnetic length}) \times (\text{pole strength})$, so we can find the strength of the poles by dividing M by the magnetic length.

Comparing values of H . With a short bar magnet and a magnetometer we could easily compare the values of H .

¹ Reading both ends of the needle and taking the mean. In accurate work, we should now repeat with the magnet reversed for end, and then take the same four readings again but on the other side of the magnet. θ is taken as the mean of all eight readings.

two different places. All we have to do is to put the bar magnet at the same distance from the needle on both occasions, and read the deflections, say θ_1 in the first case and θ_2 in the second.

Now the field produced at the centre of the compass needle is given by $F = H \tan \theta_1$ and since the same bar magnet was used in each case, at the same distance, it must have produced the same field. If on the first occasion H had the particular value H_1 , and on the second occasion the value H_2 , then $F = H_1 \tan \theta_1$ and $F = H_2 \tan \theta_2$.

$$\therefore H_1 \tan \theta_1 = H_2 \tan \theta_2.$$

$$\therefore \frac{H_1}{H_2} = \frac{\tan \theta_2}{\tan \theta_1}.$$

we know the actual value of H_1 , we can obviously calculate that of H_2 .

Magnetic Moment by the Neutral Point Method. When studying fields of force in our preceding chapter we came across various examples of *neutral points* or *null points*. Let us look again at the particular case of the neutral points observed when a bar magnet is placed with its south pole pointing north. Call them X_1 and X_2 (Fig. 32/7).

A free north pole at X_1 would be pulled northwards by the earth's 'north' magnetic pole (which is really south-seeking). From the magnet it experiences a pull southwards from S and a push northwards from N (like poles repel). N, however, is more distant, so the magnet's pull southwards is greater. The balance exerts a pull. It is because this pull southwards is equal to the earth's pull northwards that X_1 is neutral or null—a point subject to no resultant magnetic force. You can easily work out the position of X_2 for yourself.

Now suppose we could increase the strength of NS, while keeping it in position. How would the position of X_1 be affected?

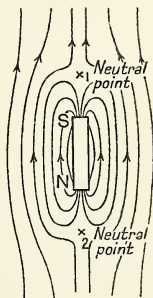


FIG. 32/7

The earth's pull on our free north pole is what it was before but the magnet's pull has increased, so X_1 is no longer neutral point—there would be a definite pull towards the magnet. If, however, the distance SX_1 is gradually increased the magnet's pull at X_1 becomes less and less, and a position of neutrality is soon reached once more. In short, *for a magnet of given length in the position we are discussing, the greater the pole strength, the more distant is the neutral point.*

Now let us go back—the magnet at its old strength and N in its original position. Remove SN and replace it by a long magnet of the same pole strength, S being in the same position as before (N , of course, will now be further 'down'). Again how will the position of the neutral point X_1 be affected?

The earth's pull northwards is unchanged. S pulls southwards with the same force as before, but N repels with less force because it is farther away. On balance, then, the magnet is pulling southwards with greater force than before and so the neutral position is once more at a greater distance from S . Similarly for the other neutral point. *For a magnet of given pole strength in the position we are discussing, the greater the length, the more distant is the neutral point.*

Since the position of the neutral points depends on both the pole strength of the magnet and on its length, we are not surprised to find that it depends on the product of these—on the *moment* of the magnet. We can, in fact, find the moment of the magnet by determining the neutral point. We shall first find the pole strength.

A long line is ruled in a north-south direction (Fig. 32/8), and a good bar magnet is placed on it, south pole pointing north.

A sensitive compass needle is moved along the line until a position is reached in which it sets east and west. The distance of the centre of the needle from S , the south pole of the magnet, is observed (S is approximately $1/12$ th of the length of the bar magnet from one end). This gives the position of X_1 , and X_2 is obtained in a similar way.

¹ This can be done directly, i.e. without first finding the pole strength. If $2l$ is the length of the magnet and $2d$ the distance between the neutral points, it can easily be shown that $M = \frac{H(d^2 - l^2)}{2d}$.

Record the distances a , b , c and d shown in Fig. 32/8.

If we imagine a unit north pole at X_1 , the force it experiences from the magnet is $\frac{m}{a^2} - \frac{m}{b^2}$ where m is the pole strength. This must be equal to H (say 0.18)

$$\therefore \frac{m}{a^2} - \frac{m}{b^2} = 0.18$$

Knowing a and b , we can calculate m . Similarly at X_2 , $\frac{m}{c^2} - \frac{m}{d^2} = 0.18$, giving us another value of m . We take the mean of the two values.

Multiplying our mean value of m by the magnetic length of the magnet, we obtain the moment.

Here is a typical result:—

Length of magnet was 10.2 cm., so the poles would be approximately 0.85 cm. from the end.

$$a = 21.9 \text{ cm.} \quad b = 30.3 \text{ cm.}$$

$$c = 22.1 \text{ cm.} \quad d = 30.5 \text{ cm.}$$

$$\frac{m}{21.9^2} - \frac{m}{30.3^2} = 0.18$$

$$\text{This gives} \quad m = 181 \text{ webers.}$$

$$\frac{m}{22.1^2} - \frac{m}{30.5^2} = 0.18$$

$$\text{This gives} \quad m = 185 \text{ webers}$$

$$\text{Mean value} \quad m = 183 \text{ webers.}$$

$$\text{Magnetic length} = 10.2 - 1.7 = 8.5 \text{ cm.}$$

$$\therefore \text{moment of magnet} = 183 \times 8.5 = 1556 \text{ c.g.s. units.}$$

So far we have assumed that the value of H is known, and we have been able to find the pole strength and moment of the magnet. Conversely, if the moment of the magnet is known, we can find H as in the following example:—

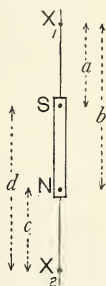


FIG. 32/8

Example 2. The magnetic length of a bar magnet is 10 cm., and its moment is 2000 c.g.s. units. When the magnet is placed in the magnetic meridian with its south pole pointing north, a neutral point is found 28 cm. from the centre of the magnet. What value does this give for H ?

Referring to Fig. 32/8, a would be 23 cm. (28 cm. is the distance from the neutral point to the centre of the magnet), and b would be 33 cm. Also m the pole strength

$$= \frac{\text{moment}}{\text{magnetic length}} = \frac{2000}{10} = 200 \text{ weber}$$

$$\therefore \frac{200}{23^2} - \frac{200}{33^2} = H$$

$$\text{giving } H = 0.19.$$

The Earth's Magnetic Field. A quantitative record of the earth's magnetic field at any place will, of course, first of all mention its direction. This is completely defined by stating the declination (i.e. so many degrees east of west of the geographical meridian) and the inclination (angle of dip). In Fig. 32/9 we will suppose that AD represents the direction of the earth's magnetic force.

We have seen how to measure the horizontal component of this force, commonly called H . In Fig. 32/9 suppose this is represented in direction and magnitude by AB .

From B draw a vertical line BC meeting AD in C . Then BC represents the vertical component of the earth's magnetic field, while AC represents the total field. In Fig. 32/9 the parallelogram $ABCE$ has been completed, and it should be clear that the total field AC is the resultant of a horizontal component AB and a vertical component BC ($= CE$).

If we know θ (the angle of dip) and H , we can easily calculate the vertical component V and the total field T , for $\frac{V}{H} = \tan \theta$ so $V = H \tan \theta$. Also $\frac{H}{T} = \cos \theta$, so $T = \frac{H}{\cos \theta}$.

Alternatively, we can draw AD at the correct angle, draw horizontally of length to represent H , and complete the figure. By measurement we can then obtain V and T .

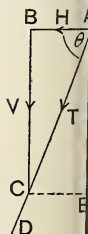


FIG. 32/9

Example 3. *If at a certain place the angle of dip is 68° and $= 0.180$ oersted, find the value of the vertical component of the earth's field, and also the total intensity.*

$$\begin{aligned} V &= H \tan \theta = 0.180 \times \tan 68^\circ \\ &= 0.180 \times 2.475 \\ &= \mathbf{0.446 \text{ oersted.}} \end{aligned}$$

$$\begin{aligned} T &= H/\cos \theta = 0.180 \div \cos 68^\circ \\ &= 0.180 \div 0.3746 \\ &= \mathbf{0.481 \text{ oersted.}} \end{aligned}$$

By drawing. Draw the triangle ABC of Fig. 32/9 making $\angle B = 68^\circ$ and $AB = 1.8$ in., to represent 0.18 oersted. BC should be found to measure 4.46 in. and AC 4.81 in. representing 0.446 oersted and 0.481 oersted respectively.)

Questions

A

In the following cases m_1 and m_2 are the strengths of two magnetic poles, and d is the distance between them. Calculate the force of repulsion between the poles.

- (i) $m_1 = 16$ c.g.s. units, $m_2 = 4$ c.g.s. units, $d = 8$ cm.
- (ii) $m_1 = 12$ „ „ $m_2 = 3$ „ „ $d = 6$ cm.
- (iii) $m_1 = 25$ „ „ $m_2 = 4$ „ „ $d = 5$ cm.

A north magnetic pole A of strength 25 c.g.s. units attracts a south pole B, 6 cm. away, with a force of 12 dynes. Find the strength of B.

Two magnets AB and CD are lying parallel to one another, with their north poles adjacent and 2 cm. apart. The magnets repel each other with a force of 27 dynes. What will be the force of repulsion if their distance apart is increased to 6 cm.?

B

State the law of force between two magnetic poles. Two magnetised knitting needles are suspended parallel to each other and 2 cm. apart with their like poles adjacent. The needles repel each other with a force of 4 dynes. How far apart must they be hung for the force to be reduced to 1 dyne? *Camb.*

2. Define *magnetic pole strength* and *magnetic field strength*.

A bar magnet 10 cm. long and of pole strength 40 units placed horizontally, in an east-west position, with its north pole pointing east. Find, from first principles, the field strength at a point P, 10 cm. east of the north pole on the prolongation of the axis.

If the horizontal component of the earth's field is 0.2 oersted, what is the resultant magnetic field at P? *Lond.*

3. State the law of force between magnetic poles.

The magnetic field due to a bar-magnet (length 20 cm., cross-sectional area 0.5 sq. cm.) is 0.16 gauss at a point end-on to the magnet and 40 cm. from its centre. Calculate (a) the magnetic pole strength on the assumption that the poles are at the ends of the magnet; (b) the average intensity of magnetisation.

State, with reasons, whether you consider this a suitable method for measuring pole-strength. *O.C.*

4. Define the magnetic moment of a magnet.

Describe how you would compare the magnetic moments of two short bar magnets.

Find the couple required to hold a bar magnet of pole strength 50 c.g.s. units and magnetic length 12.4 cm., so that it is at right angles to a magnetic field of strength 0.2 c.g.s. units. *Oxf.*

5. Two bar magnets of magnetic moment 100 and 800 units respectively, are to be placed on the arms of a magnetometer, with its arms in an East-West direction, so that there is no deflexion. What must be the ratio of the distances of the magnets from the needle? Explain your calculation. You may assume that the distances will be large compared with the lengths of the magnets. *O.C.*

6. A short bar magnet A is placed 20 cm. from the needle of a magnetometer used in the ordinary way, and causes a deflexion. A second short magnet B is brought up on the opposite side of the needle, A still remaining in place. When B is at a distance of 25 cm. from the needle the deflexion is reduced to zero. Compare the magnetic moments of A and B. *J.M.B.**

7. Define *magnetic meridian* and *magnetic dip*.

With the use of a compass needle and a short bar magnet, describe a method for comparing the horizontal components of the earth's magnetic field at two places.

Find the horizontal component of the earth's magnetic field at a place where the total intensity of the field is 0.47 gauss and the angle of dip is 60° . *O.C.*

8. What is meant by a *line of magnetic force* ?
 A ball-ended magnet is placed in the Earth's field with its axis horizontal and in the magnetic meridian, and with its south pole towards the north. Draw a diagram of the lines of magnetic force, marking the directions of the lines with arrows and indicating the positions of the neutral points.
 If with a magnet in this position neutral points exist 15 cm. from the poles and the distance between the poles is 20 cm., calculate the pole strength and the magnetic moment of the magnet. ($H = 0.18$ gauss). *Camb.*
9. A bar-magnet of length 10 cm. and magnetic moment 1300 units is placed in the magnetic meridian with its north pole pointing south. A neutral point is found at a distance 25 cm. from the centre of the magnet. Calculate the strength of the Earth's magnetic field. *O.C.**
10. A bar magnet is placed with its axis in the magnetic meridian and north pole pointing south. Draw a diagram of the lines of force which you would expect to find around the magnet. If the pole strengths of the magnet are 120 units, its length 10 cm. and a neutral point is found 30 cm. from the centre of the magnet, calculate the strength of the horizontal component of the Earth's magnetic field. *O.C.**
11. A magnet is placed on a table with its north pole pointing north. If it is 10 cm. long and has a pole strength of 50 c.g.s. units, what is the magnetic field strength at a point 10 cm. from the north pole and on the prolongation of the axis of the magnet? ($H = 0.18$ oersted.) *Lond.**
12. Two magnets, each of magnetic length 8 cm. and of pole-length 40 and 60 c.g.s. units respectively, are placed so that they lie along the same straight line with like poles 4 cm. apart. Calculate the resultant force of repulsion between the magnets. *C.W.B.**
13. A bar magnet of magnetic length 12 cm. and moment 100 c.g.s. units is placed in the meridian with its south pole pointing north. A neutral point is found 30 cm. from the centre of the magnet. Calculate the strength of the earth's magnetic field. *O.C.**
14. In the magnetic observatory at Abinger (Surrey), for the year 1948, the following observations were recorded : Declination 35° W. Dip $66^\circ 46'$. Horizontal component 0.1858 dynes per unit pole.
 Explain the meaning of each of these terms, illustrating your answer by diagrams. Calculate the vertical component and the total magnetic intensity.

CHAPTER 33

ELECTROSTATICS

A. Electrification by Friction

THE article on 'Electrostatics' in the *Encyclopaedia Britannica* occupies about seventeen columns. If there had been such a publication in the days of the ancient Greeks it would have occupied about seventeen lines, merely stating that when *elektron* (amber) was rubbed with cloth it acquired the curious property of attracting very light particles, such as bits of thread. The Greek schoolboy could acquire a complete knowledge of the science of electricity on very easy terms!

In electrostatics we are concerned with electricity at rest, as distinct from current electricity, a term which explains

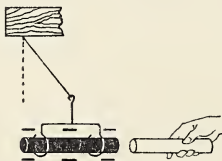


FIG. 33/1

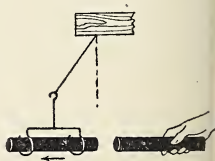


FIG. 33/2

itself. If we regard the one as resembling a lake, the other is like a stream.

In these days we do not experiment with amber, but we obtain similar effects by rubbing a vulcanite fountain pen with cloth, or a glass rod with silk. The vulcanite or glass will then attract light bodies such as bits of paper, fragments of cork, etc. Thus *an electrified body will attract an unelectrified one*. It follows from Newton's Third Law of Motion that an unelectrified body will attract an electrified one, but we can easily prove this directly. We support an electrified vulcanite rod in a stirrup of copper wire suspended from a silk thread (Fig. 33/1), and bring near it various unelectrified bodies in turn—a book, a glass rod, one's hand, etc. In each case the vulcanite is attracted.

Figs. 33/2, 3 and 4 show respectively

- (i) an electrified vulcanite rod brought near another electrified vulcanite rod;
- (ii) an electrified glass rod brought near another electrified glass rod;
- iii) an electrified glass rod brought near an electrified vulcanite rod.

In cases (i) and (ii) we have repulsion, but in (iii) we have attraction. We conclude that *like charges repel one another, while unlike charges attract*.

Experiments such as the foregoing were made by Dr. William Gilbert of Colchester, late in the sixteenth century. He found that he was unable to obtain any results with certain

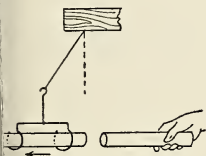


FIG. 33/3

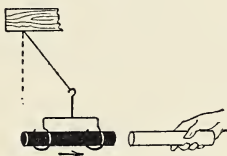


FIG. 33/4

ones, particularly if they were made of metal. He wrongly concluded that such rods could not be electrified, and called them *non-electrics*, substances such as glass being *electrics*. We now know that his 'non-electrics' were really *conductors* of electricity, so that the charges leaked away as fast as they were produced. If a brass rod is mounted on an insulating stand (e.g. one made of vulcanite), it can be electrified by friction in the usual way.

Modern Theory of Electrification. Franklin, a famous eighteenth-century experimenter, believed that whatever electricity there might be, there was only one kind of it, and all bodies possessed it. When a body held its normal supply it was 'uncharged'—had no power of attracting bits of paper. In certain circumstances, however—when rubbed, for instance—it might gain electricity, in which case it was 'positively' charged. Or it might lose it, and then its charge was 'negative.' This simple theory explained the known

facts very well. During the last fifty years, however, our knowledge of the nature of the atom has greatly increased and this has led to the following view of the nature of electricity.

Probably you know from your studies in chemistry that though a small portion of an element—lead, for instance—can be divided into halves, quarters, eighths, sixteenths and so on, this subdivision cannot go on for ever, even if we had a knife of infinite sharpness and sight of infinite keenness. We come at last to the atom.¹

The atom is inconceivably small, something like 10^{-8} cm. in diameter. This conveys very little meaning to us. Imagine an enormous goods-train stretching right round the world at the equator—25,000 miles—every waggon in being loaded up with wheat. The number of grains of wheat would be far less than the number of molecules of hydrogen in a cubic centimetre (the molecule of hydrogen consists of two atoms); and remember that a cubic centimetre is much less than a thimbleful, and that the molecules are by no means closely packed, either!

But small as is the atom, it has a *structure*, and includes parts which are very much smaller than itself. These parts are of three different kinds—electrons, protons and neutrons. All the different atoms—lead, hydrogen, sulphur, etc.—contain these parts. Different numbers of them, and different arrangements, give the different atoms (just as different numbers and arrangements of bricks, tiles, beams, etc., give us different houses).

The protons and neutrons are bunched together to form what is called the *nucleus* of the atom—the comparatively heavy, 'solid' centre. It would seem that the electrons revolve round the nucleus somewhat as planets revolve round the sun. In fact, the solar system is something like an atom on an immense scale, the sun representing the nucleus (made up of protons and neutrons) and the planets—Mercury, Venus, the Earth, Mars, etc.—representing the electrons. And as the mass of each planet is only a tiny fraction of the sun's

¹ Strictly speaking, to a small group of atoms known as the *molecule*. Sometimes there is only one atom in the molecule, in which case the atom and the molecule are, of course, the same thing.

mass, so the mass of the electron is only a tiny fraction of the mass of the nucleus. In fact, the mass of an electron is only about the 1800th part of the mass of one of the neutrons or protons which make up the nucleus.

It will help to clear up our ideas if we examine Fig. 33/5 for a minute. It represents an atom of iron, with a heavy nucleus made up of protons and neutrons, and a number of electrons revolving in various orbits about the nucleus. For simplicity the diagram shows these orbits as a series of concentric circles in one plane. It is certain that they are nothing like this, but that is a point which need not concern us here.

It will be clear that an atom is a very *porous* structure. It is actually far more porous than the diagram would suggest, on the scale to which the nucleus is drawn, the outer orbits should have a diameter of over 30 ft. ! Scientists have

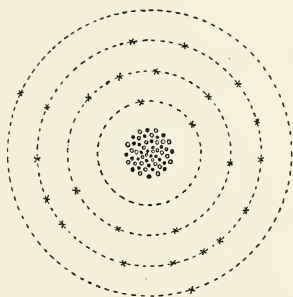


FIG. 33/5

usually succeeded in bombarding the atom (using neutrons as their 'bullets') and they find that for every bullet that hits the central nucleus, thousands go clean through.

It is certain that the electron is negatively charged. It appears, in fact, that the electron is a negative charge, and that negative electricity is, so to speak, 'granular,' consisting of electrons, every one of which represents an equal charge.

Now, the atom as a whole, nucleus plus electrons, is electrically neutral. Since the electrons are negative charges, it follows that the nucleus must contain a balancing positive charge. The latter is contributed by the protons. The atom of iron, for instance, contains 26 protons¹ (represented in Fig. 33/5 by small white circles), and the 26 positive charges

And 30 neutrons. As regards mass, 1 neutron = 1 proton (approx.) so this is practically the chemical unit of atomic weight. Thus, neglecting the very small weight of the electrons, atomic weight of iron = $26 + 30 = 56$.

are balanced by the 26 negative charges represented by the electrons.

In certain circumstances the atom (neutral) may lose one or more of its negatively charged electrons. The remaining part of the atom will then evidently be positively charged. It is the better conductors which tend to lose electrons. It is very difficult to detach electrons from an insulator.

With these ideas in mind, let us try to explain what happens when (for instance) vulcanite is rubbed with fur or flannel. Some of the molecules (i.e. groups of atoms) of the fur lose their electrons, which become attached to the surface of the vulcanite. The latter will therefore be negatively charged, leaving the fur positively charged. One has simply gained what the other has lost, and so the two *taken together* are still electrically neutral, as we saw in the experiment described on p. 328 of our earlier book.

We can now give a reasonable explanation of the well-known fact that flame has a discharging effect on a charged electroscope, vulcanite rod, etc. We must remember that the molecules of a hot gas are in a state of specially violent vibration. It seems likely that, as a result, some of them lose electrons (thereby becoming positively charged), while others gain them, becoming negatively charged. The molecules having a charge opposite to that on the vulcanite will quickly be attracted to it, and will neutralise its charge, or we may say, those molecules which have lost an electron will combine with the free electrons on the surface of the vulcanite, thus discharging it.

B. Electrification by Induction

In our earlier work we made much use of the gold-leaf electroscope. The diagram (Fig. 33/6) will serve to recall it. Notice that the brass disc A, the brass rod BC and the gold leaves G, G' are all in electrical connection with each other, but are *insulated* (by suitable material inside the cork) from the outer part of the apparatus. If we beat the disc with flannel, it receives a positive charge by friction. This charge is conducted to the leaves, which then repel one another because they have received 'like' charges. We have charged our electroscope positively, by friction.

But we can give it a positive charge in quite another way. We bring near it a *negatively* charged rod (vulcanite rubbed with flannel). The leaves at once diverge. We may suppose that the vulcanite rod has attracted a positive charge to the brass cap, repelling an equal negative charge to the leaves (Fig. 33/7a). We now touch the cap for a moment. The negative charge escapes to earth through the body, and the leaves collapse. The positive charge on the cap, however, is still 'held' by the negative charge on the vulcanite (Fig. 33/7b). We now remove the finger, and then the rod. The positive charge distributes itself over cap, rod and leaves, and the latter diverge once more, this time with a positive charge. We have charged our electroscope by *induction* (Fig. 33/7c).

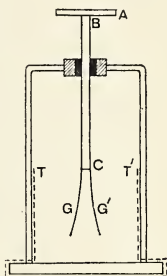


FIG. 33/6

One of the many uses of the electroscope is to find (i) whether a body possesses an electric charge or not, and (ii) in the former case, the nature of the charge. To find whether a body is charged, we bring it near the

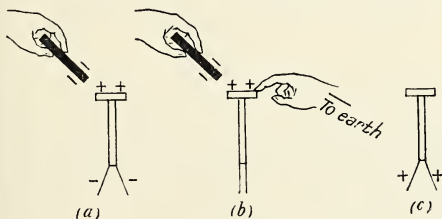


FIG. 33/7

of an uncharged electroscope. If the body is charged, the leaves will diverge. (Instead of bringing the body itself, one may use a small metal disc mounted on an insulated stand—a *proof plane*—to convey a sample of the charge, if any, on the body.)

Suppose the body is found to be charged, and we wish to determine whether the charge is positive or negative. We first

charge the electroscope by induction—say positively, already described. We now gradually bring either the body itself, or a proof plane which has been in contact with it, near the cap of the instrument. If the leaves diverge more widely the charge was obviously positive. If they gradually collapse it was negative.

Faraday's ice-pail experiment. A very instructive series of experiments on Induction was carried out by Faraday using a pewter ice-pail—he used anything that was handy. We repeat his work using a rather tall cocoa-tin or a calorimeter.

I. The tin is placed on the cap of an electroscope, and a metal ball charged (say) positively, and suspended by a

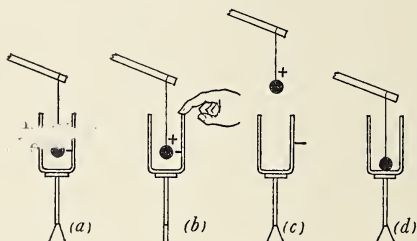


FIG. 33/8

thread, is lowered well inside the tin, without touching the sides.

The leaves diverge (Fig. 33/8a). On bringing a positively charged glass rod near the electroscope, the divergence is increased, showing that the charge on the electroscope is positive. This is what we should expect from our knowledge of induction.

On moving the metal ball about, inside the can, but not touching it, and not very near the mouth, the leaves are quite unaffected. Evidently the magnitude of the induced charge does not depend on the position of the charged body inside the hollow vessel.

On withdrawing the ball, the leaves collapse. The electroscope when tested is found to have retained its charge.

II. The ball is again introduced, and of course the leaves

diverge once more, but they collapse when the can is touched with the finger (Fig. 33/8b). The finger and then the ball are now withdrawn. The leaves diverge to the same extent as before, but the charge on the electroscope is now found to be negative. The ball still retains its charge (Fig. 33/8c).

When the can was touched, the induced positive charge escaped to earth, and only the induced negative charge remained. This must have been *equal* to the induced positive charge, for the leaves diverged to the same extent.

III. We now discharge the electroscope, so as to be exactly where we were at the beginning of the experiment. The ball is once more introduced, making the leaves diverge (Fig. 33/8d), but the ball is now made to touch the inside of the can. The divergence of the leaves remains unaffected. On withdrawing the ball it is found to be discharged. The leaves still remain apart, and the can and electroscope are now found to be positively charged.

Since the ball was found to be discharged the induced negative charge on the inside of the can must have been equal to the positive charge on the ball. Summing up—

from II, induced negative charge = induced positive charge.
 from III, induced negative charge = inducing positive charge
 on ball.

Hence—*when a charged body is introduced into the inside of a closed hollow conductor, each of the induced charges is equal in magnitude to the inducing charge; and therefore when the charged body is made to touch the inside, the entire charge is transferred to the outside.*

Now, when we bring a charged body nearer and nearer to the electroscope, the leaves diverge increasingly. We might be tempted to think that in this case the induced charge was *less* than the inducing charge, but was increasing as the latter approached. Actually the *total* induced charge is always exactly equal to the inducing charge, but not all of it is on the electroscope. Some of it is on the walls of the room, etc. As the charged body is brought nearer, a larger proportion of the induced charge is concentrated on the electroscope, making it show a greater divergence.

Electric Screening. It seemed to Faraday that as the charge on a hollow conductor was entirely on the outside, the inside should be free from the action of electric forces. He put this idea to a very severe test. He made a box about 6 ft each way and covered the outside of it with conducting material such as tin-foil, wire gauze, etc. Inside the box he had electroscopes and other delicate instruments. The box was placed on insulating supports, and charged with electric machines to such an extent that an assistant was able to obtain long sparks from the outside, but the instruments inside were quite unaffected.

A much simpler but quite effective experiment on 'electric screening' can be carried out with an electroscope and a lighted taper. The electroscope is charged, and a lighted taper is placed near the cap. The electroscope is quickly discharged (p. 404). It is re-charged, but this time a wire gauze is held between the lighted taper and the cap. The leaves now remain open.

Lines of Electric Force. When studying magnetism we gave some attention to *lines of magnetic force*—the lines along which free north poles would travel under the action of magnetic forces. We found that such lines could be obtained by the use of iron filings, or of a compass needle. But we found also that they could be constructed, though very laboriously, by using our knowledge of attractions and repulsions, and of the inverse square law. Let us see how we can make use of similar knowledge in electricity¹ to plot *lines of electric force*, such a line being *the line along which a free positive charge of electricity would move under the action of electric force*.

To begin with a very simple case, if the 'electric forces' are due to a positive charge P, the lines of electric force will be as shown in Fig. 33/9a. If P is a charged *sphere*, we may suppose the charge to be concentrated at its centre, and the lines of force will run in the same way (Fig. 33/9b).

If the lines of force are due to a *negative* charge P, it is clear that a free positive charge would be *attracted*, so the lines point in the opposite direction (Figs. 33/10a and 10b).

Next we come to the much harder case of when *two* charges

¹ It can be shown that the inverse square law also applies to electrostatic charges.

re concerned, a positive charge at P and an equal negative charge at N (Fig. 33/11), both, of course, *fixed*. If another

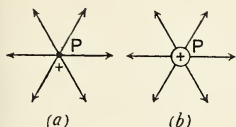


FIG. 33/9

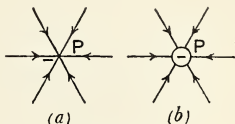


FIG. 33/10

positive charge, *free to move*, is at p , which way would it travel? It is repelled by P in the direction pa , and attracted by N in the direction pc . To represent these forces we must make the ratio pa/pc equal to $(pN)^2/(pP)^2$ (law of inverse squares). For instance, if $pP = 1$ unit and $pN = 2$ units, we should have to make pa 4 times pc . Having drawn pa and pc of the correct relative lengths, we complete the parallelogram $pabc$; and pb is the direction in which p would *begin* to move. After it had moved a very little way, its direction would alter, because it would be farther from P and nearer to N. Its complete path would be somewhat as shown by the dotted line PpN . If a number of such lines

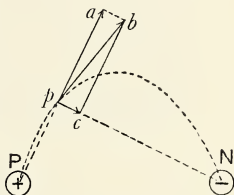


FIG. 33/11

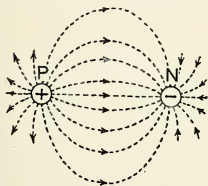


FIG. 33/12

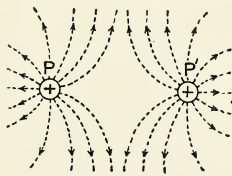


FIG. 33/13

we drawn, we obtain Fig. 33/12. On the other hand, if P and P' are both positive charges, the lines of force run as in Fig. 33/13.

There are some important characteristics of lines of electric force, which we will briefly summarise.

- (1) Starting from a positively charged body, they always end on one with a negative charge. This may be on a body with a *free* negative charge (Fig. 33/12). Failing this, they end on a body with an *equal induced* negative charge. This charge might be on the wall of the room, the ceiling, etc. In fact, the charges at the two ends of a line of force are always equal and opposite.
- (2) They always enter or leave the surface of a conductor at right angles.
- (3) They tend to shorten themselves, like cords of stretched elastic. In Fig. 33/12 this would cause P and I to close in on one another. We may regard it as the reason why 'unlike charges attract one another.'
- (4) Lines of force running near to one another, and in the same direction, repel one another. This would cause P and P' in Fig. 33/13 to move apart. 'Like charges repel one another.'

C. Electrostatic Units

In the preceding pages we have frequently dealt with charges of electricity produced by friction. We might perhaps have described them by such terms as 'strong' or 'weak,' but we had no means of accurately *measuring* them because we had no *unit*.

At first sight it seems rather hopeless to have to decide upon a unit of electricity, but we have a key to the problem—we know that one charged body will repel another that is similarly charged. They repel one another with a certain *force*, and we can measure force in dynes. The force, of course, depends on the distance between the two charges, and later on we shall find that it varies with the medium that happens to separate the two charged bodies (usually but not always, air).

We now have all the material necessary for defining our unit. **Unit electric charge** is *such a charge that, if placed in vacuo at a distance of 1 cm. from a similar charge, it will cause*

the latter to be repelled with a force of 1 dyne. (The force in air differs very little indeed from the force *in vacuo*.) We notice a close resemblance to the definition of unit magnetic pole on p. 385.

As already mentioned, it can be shown that the force between two electric charges follows the inverse square law—i.e. it is inversely proportional to the square of the distance. Expressed in a formula, if two charges q and q' are separated in air¹ by a distance of d cm., the force between them is $\frac{qq'}{d^2}$.

Example 1. Two small spheres carry electric charges of 20 and 30 units respectively, and the distance between their centres is 10 cm. Find the force of repulsion between them.

The effect is the same as though the charges were concentrated at the centres of the spheres, so the required force = $\frac{20 \times 30}{10^2} = 6$ dynes.

Electric Field. The *electric field at a given point* is the force in dynes that would be experienced by unit electric charge placed at that point.

Example 2. Find the strength of the electric field at a point between the centres of the spheres in Example 1, and 10 cm. from the centre of the second sphere.

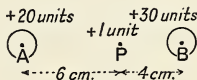


FIG. 33/14

Suppose A is the centre of the first sphere, B of the second. P be the point. Then $AP = 6$ cm. and $BP = 4$ cm. If we imagine unit electric charge placed at P, then

$$\text{force due to A} = \frac{20 \times 1}{6^2} = 0.56 \text{ dynes.}$$

$$\text{,, ,, B} = \frac{30 \times 1}{4^2} = 1.88 \text{ ,,}$$

These forces are in opposite directions, the first acting in the direction PB and the second in the direction PA. The resultant field is therefore one of $1.88 - 0.56$ or 1.32 dynes in the direction PA.

If these are separated by some other medium, the force is qq'/Kd^2 , where K is the dielectric constant (p. 418).

D. Potential

We all know that if two quantities of water are connected say by a pipe, there will be a flow from the water at the higher level to that at the lower. If there is no difference of level no flow will take place. It is not a question of *quantity*. In hundreds of cases we have water flowing from a lake to the sea (a river-bed in this case taking the place of the pipe) although the quantity of water in the lake is much less than that in the sea.

Now, just as difference of *level* causes flow of water (and we might add, just as difference of *temperature* causes flow of heat), there is a difference of something which causes flow of electricity. This something is known as *potential*—its difference of potential (often written as 'P.D.') which causes flow of electricity.

Zero Potential. An ordnance map gives the heights at various places 'above sea-level.' Sea-level is a very convenient 'zero,' because, although the sea is constantly suffering losses and gains, these are so small compared with the total vast bulk as to make no perceptible difference to the level.

In the same way, the electric charges which are constantly flowing into and out of the earth are so small, compared with the earth's vast store of electricity, as to leave it unaffected for practical purposes. We can no more raise or lower the earth's electric level (i.e. its 'potential') than we can raise or lower the level of the water in the ocean.

For this reason, the earth is said to be at *zero potential*. If, when a charged body is connected to earth (A in Fig. 33/15) positive electricity flows from it to the earth, the potential of that body is said to be positive. If positive electricity flows in the opposite direction (B in the figure), the potential of the body is negative. If there is no current in either direction, the potential of the body is zero.

P.D. indicated by Electroscope. We have often used a gold-leaf electroscope to indicate the presence of an electric

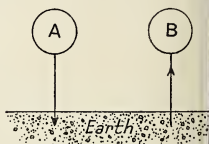


FIG. 33/15

large, or to tell us which of two charges is the greater. Indirectly, as we shall see later, the electroscope does give us information about electric charges, but what it *directly* indicates is difference of potential, as the following experiment will show.

Place the electroscope on an insulating support—say a slab of paraffin wax. Take a piece of copper wire about a foot long and pass it through a piece of rubber, which will serve as an insulating handle. Connect the wire to the cap of the electroscope (Fig. 33/16).

Now charge the electroscope by touching the cap with a proof-plane which has itself been charged by contact with the brass disc of an electrophorus (*Junior Physics* p. 321). If necessary, increase the charge by repeating the operation with the disc once or twice, until the leaves show a good divergence.

Next, holding the copper wire by the rubber handle, make its lower end touch the tin foil T. The leaves at once collapse. Yet the electroscope cannot have lost its charge, for the wax slab is insulating it from the earth. We can confirm this point (that the electroscope still has a charge) by bringing a second electroscope near it. The leaves of the latter diverge.

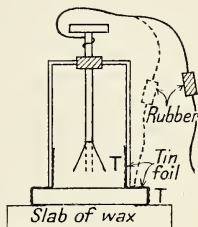


FIG. 33/16

Now, when we made the wire touch the tin-foil, the latter would acquire the same potential as the cap and leaves. It is clear, then, that when there is no difference of potential between leaves and tin-foil, there is no divergence of the leaves; to put the case the other way round, divergence of the leaves indicates a difference of potential between leaves and tin-foil.

We can vary the experiment by first of all giving a charge to the tin-foil (the electroscope being insulated as before).

The leaves diverge (though they have as yet received no charge). On connecting cap and tin-foil, the leaves collapse before.

But although what the electroscope *directly* indicates is potential difference between leaves and tin-foil, indirectly it does indicate the presence of a charge of electricity on the leaves. For, as ordinarily used (we shall assume this through-

out the following little argument), the tin-foil is earthed—i.e. it is at zero potential. Therefore, since divergence indicates a difference of potential between leaves and tin foil, the leaves themselves must have a potential that is not zero. But to have a potential, they must have received charge¹ (just as to acquire a temperature, a body must receive heat). Thus divergence of leaves does indicate charge, and the greater the charge, the greater the divergence.

This last point is easily proved experimentally. We take two exactly similar electroscopes. We charge the brass plate of one with an electrophorus, and by means of a proof-plane transfer what we will call one 'unit' to the first electroscope. To the second, we transfer two or three units. Clearly the second electroscope has received the greater charge, and it also shows the greater divergence.

Potential of a Conductor. A moment's thought will satisfy us that the potential at all points of a charged conductor is the same, e.g. that shown in Fig. 33/17—must be the same. If,

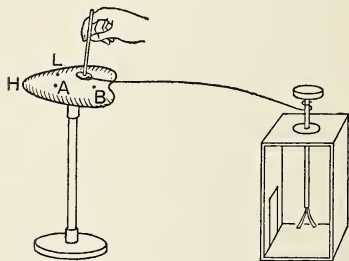


FIG. 33/17

for instance, the potential at A were higher than at B, electricity would flow from A to B until the potential at both points was the same. We can compare this with the fact that points on a sheet of calm water are at the same level.

To prove the statement experimentally, we may use

¹ Or alternatively they must be in the neighbourhood of some body which is not at zero potential—an alternative which in this case is ruled out.

paratus shown in Fig. 33/17. The proof plane is connected by a wire with the cap of the (earthed) electroscope. No matter to what part of the conductor we apply the proof-plane, the divergence is the same, and the divergence, as we have seen, is a measure of the potential.

Potential the Result of Work Done. To raise the level of water, work has to be done against the force of gravity. We then get an example in a country house, where water may be pumped from a well to a tank in the attic.

Correspondingly, when we do work on a charge of electricity, we raise its 'electric level'—i.e. its potential. Suppose, for instance, that at A and B (Fig. 33/18) there are two small bodies, one carrying a charge Q , the other a charge q (both positive). Between the two there is,

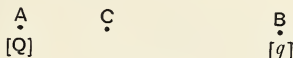


FIG. 33/18

of course, a force of repulsion.

1. Suppose we bring q

to a point C, nearer to A. To do so, we have to *do work* against this force of repulsion, and at C the charge will be at a higher electric level than when it was at B.

When released, the charged body will tend to move from C to B. That agrees very well with what we already understand—that electricity tends to move from a point of higher potential to one of lower potential.

Potential is a quantity which can be accurately measured. Electricity is supplied to your house at a certain *voltage*—perhaps: i.e. its potential exceeds that of the earth by so many volts, the earth's potential being regarded as zero. The volt is a practical unit of potential. Let us give it a little more attention.

Go back to Fig. 33/18, where there is a charge Q at A. Suppose B is very far removed from A—say a mile—and at B there is *unit* charge (defined on p. 410). We begin to move this charge from B to C.

For much the greater part of the journey we meet with no appreciable resistance. As we get rather near to C we begin to feel the force of repulsion due to the charge at A—a force which can be measured in *dynes*. To overcome this resistance

we have to do work which is measured in *ergs*. The total amount of work which we do in bringing our unit charge to the point C, is the potential at C, and if this amount of work is 1 erg, the potential at C is unity.

To be perfectly correct we must make the original distance not a mile, but infinity, and we can now give a formal definition of the E.S.¹ unit of potential. *In the E.S. system of units if the work done in bringing unit charge from infinity to a certain point is one erg, then the potential at that point is unity.*

The unit has received no special name—it is simply ‘the E.S. unit of potential.’ The *volt* is $\frac{1}{300}$ th part of this unit.

From the definition given above you can probably prove for yourself that the difference of potential between two points is equal to the work done in moving unit charge from one point to the other.

Capacity. Suppose A is a small conductor and B a similar but much larger one, and we give equal charges to each; it can easily be shown by experiment that the potential of A rises more than that of B.

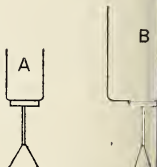


FIG. 33/19

It is more convenient for experimental purposes (though not essential) if A and B are hollow tins which are placed on similar electroscopes (Fig. 33/19). We can then transfer the *whole* of the charge from a proof-plane to the tins. As our ‘store’ of electricity, we might use the charged brass plate of an electrophorus, and make, say, three transfers to A and three to B. We give B the first transfer and then alternate them, so that if there is some falling off in the amount transferred, one shall not have all the strongest charges and the other all the weaker ones.

The ‘B’ electroscope shows less divergence than the ‘A’ electroscope, indicating that its potential has not risen so much. Because it would need a larger charge to make the potential of B as much as that of A, we say that it has a larger *electrical capacity*, a term which may be defined as the quantity of electricity needed to produce unit increase of potential. Thus, if the potential of a conductor rises by V (volts) when

¹ E.S. (also e.s.) = electrostatic.

receives a charge of Q (coulombs), its capacity C would be $\div V$, because this quantity would make its potential rise *one volt*. Capacity is measured in *farads*. In short, $C = Q/V$, where C is measured in farads, Q in coulombs and V in volts. For practical purposes, a unit called the *microfarad* (=one-millionth of a farad) is commonly employed.

The Condenser. The condenser is an arrangement by means of which the capacity of a given conductor may be greatly increased. The underlying principles may be understood from the following experiment.

i) Mount a rectangular zinc plate A on an insulating base, and connect it with an electroscope (at this stage B is not supposed to be present). Charge A by means of an electrophorus, using the leaves of the electroscope to diverge.

Now take a similar zinc plate. Hold it by means of rubber bands to insulate it, and make one vertical edge touch a vertical edge of A , but with a little overlap. For practical purposes we have *enlarged* A . The electroscope shows a reduced divergence.

ii) Take away the supplementary plate, thus reducing A to its original size. Now bring up B , a plate similar to A , but earthed. It is held parallel to A , but is placed that at first only a small portion of its area is opposite A . Slide it along (guided by a chalk line drawn on the table), so that more and more of its area is opposite to A . The divergence grows steadily less.

iii) Next bring B nearer and nearer to A . The thinner the layer of air between the plates, the less the divergence.

iv) Lastly, with A and B at a fixed distance apart, interpose between them, one at a time, slabs of vulcanite, paraffin wax, glass, ebonite, etc., taking care not to earth A by accidental contact. Observe what happens when *water* is used. (Interpose one or two glass bottles containing water, broad side parallel to the plates.)

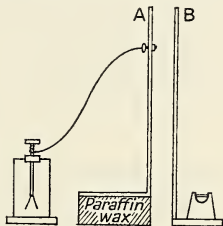


FIG. 33/20

Plate A is known as the *collecting plate*, B as the *condensing plate*.

The insulating medium separating the plates is called the *dielectric* (air in the early parts of the experiments, later changed as far as possible to vulcanite, paraffin wax, etc.). From our results we may draw the following conclusions.

The capacity of a condenser may be increased by

- (i) increasing the area of the collecting plate;
- (ii) increasing the effective area of the condensing plate ('effective area' means that portion of the condensing plate which is opposite the collecting plate);
- (iii) reducing the thickness of the dielectric;¹
- (iv) using a dielectric other than air.

More exact experiments have shown that the capacity actually *proportional* to the area of the collecting plate and to the effective area of the condensing plate, and *inversely proportional* to the thickness of the dielectric.

With regard to (iv), it has been found that the substitution of another dielectric for air (strictly, for vacuum) causes the capacity to be increased in a definite ratio—e.g. using vulcanite the capacity is increased 3.15 times, paraffin wax 2.3 times, and water actually 80 times. This ratio is known as the *specific inductive capacity*, or *dielectric constant* (usually represented by K).

i.e. *dielectric constant of a given medium*

$$= \frac{\text{capacity with that medium as dielectric}}{\text{capacity with air as dielectric}}$$

Condensers in Actual Use. The condenser was discovered by accident. In 1746 a Dutch physicist, Pieter van Musschenbroek, of Leyden, thought he would like to electrify some water contained in a bottle. To do it, he attached a wire to the prime conductor of an electrical machine, the wire dipping into the water. He held the bottle in his hand. After the machine had been in action for some time, he proceeded to move the wire with his other hand. The result must have been very startling, for he declared that 'he would not take another such shock for the kingdom of France.'

We can now understand what had happened. When very high voltages are concerned water is a conductor. The capacity of the water (plus wire) was greatly increased by the presence of the man's hand only a few millimetres away (the thickness of the glass). It would therefore take a much

¹ A very simple example of this is seen when the hand is held near the cap of a charged electroscope. As the hand is brought nearer, the leaves gradually close in. The hand is, of course, acting as the condenser plate.

greater charge of electricity before its potential reached that of the prime conductor. When the man took hold of it he received this charge through his arms and shoulders, and seems to have been most unwilling to repeat the experiment.

This accident gave van Musschenbroek the idea of the Leyden jar, a modified form of the apparatus that gave him the shock.¹ In its modern form (Fig. 33/21) we have a glass jar coated inside and outside with tin-foil for about half its height, the upper half being covered with shellac varnish to improve the insulation. Through an insulating cover there passes a brass rod terminating above in a knob, and below in a chain which makes contact with the inner tin-foil surface.

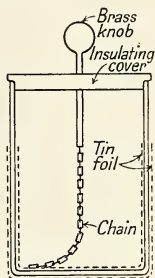


FIG. 33/21

To charge it, the jar is held in the hand (which touches the outside tin-foil) while the knob is in contact with the prime conductor of an electrical machine, say a Wimshurst.

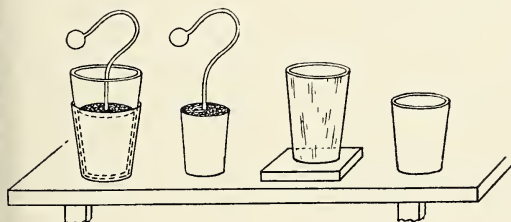


FIG. 33/22

At first sight the Leyden jar looks very unlike the arrangement discussed on p. 417. Comparing Fig. 33/20 with Fig. 33/21, however, we see that the collector plate A is repre-

It is difficult for us in these days to realise the excitement produced by the invention of the Leyden jar. Men used to make their living by going one round the country, arranging a sort of 'show' in various towns and villages, and, of course, making a charge for admission.

sented by the inner tin-foil and the condenser plate B by the outer one. The only difference is that they form cylinder instead of flat surfaces, while instead of air as dielectric we have glass. This has a dielectric constant of about 8, and the capacity of the condenser is increased to that extent.

For experimental purposes, Leyden jars with movable plates are sometimes used. One is shown in Fig. 33/22 and it can be used to show that the charge is not on the metal surfaces at all, but *on the dielectric*. To show this the condenser is first charged in the ordinary way and earth-connected by placing it on the table. The inner coating is lifted out by means of an ebonite rod, and then touched with the hand. No spark passes. The glass is removed and placed on a block of paraffin wax. The outer coating is, of course, connected with earth, so cannot contain a charge. The condenser is now re-assembled, the inner metal part being lifted by means of the ebonite rod. The inner and outer coatings are connected in the usual way by discharging tongs, when a bright spark passes!

As the charge was not in the metal coatings it must have been on the glass. In fact, if the condenser is again charged and separated as before, it is possible to obtain tiny discharges by holding a lead pencil against the glass.

Another form of condenser is shown in Fig. 33/23. Here sheets of glass are coated with tin-foil. Alternate sheets of foil are joined up to a common terminal T_1 , and those in between to T_2 . The effect is the same as if we had two very large sheets of tin separated by glass. Sometimes mica (or in cheaper instruments, wax paper) is used instead of glass, the sheets of tin in this case not being actually coated on to the mica or paper.



FIG. 33/23

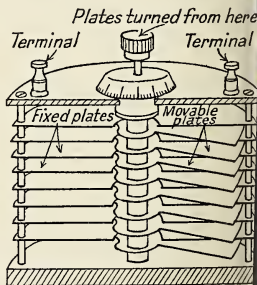


FIG. 33/24

Often, as in a wireless set, we need a condenser of which we can vary the capacity. Here we make use of the fact that the capacity is proportional to the effective area of the plates. Fig. 33/24 will almost explain itself. We have a fixed set of metal plates, and another set which can be rotated so as to move in or out of the fixed set. The dielectric in this case is usually air, but in some modern patterns, mica sheets are attached to the plates, and act as the dielectric.

Condensers are used extensively in wireless telegraphy and telephony. They also generally form part of an induction coil (p. 501), which in turn has various uses (e.g. for 'coil ignition' in a motor car). They play an important part in the production of X-rays, and have received a number of other important applications.

Questions

Gilbert believed that solid substances could be classified *electrics* and *non-electrics*. What did he mean by these terms? Do you think his classification was right? If not, explain how he came to make the mistake, and give a better classification.

Give some account of the modern theory of electrification.

Describe a gold-leaf electroscope. How would you use it to show the difference in insulating properties of silk, cotton, paper and mica?

How could you show that equal and opposite charges are produced by friction?

How would you use a gold-leaf electroscope (i) to find whether an insulated metal ball was electrically charged or not, (ii) in the former case, to find whether the charge was positive or negative?

A metal can stands on the plate of an uncharged electroscope, a charged metal ball, suspended by a silk thread, is lowered into the can.

Describe and illustrate by diagrams what happens in the following sequence of independent operations: (a) the ball is lowered until well inside the can and is then withdrawn, (b) the ball is lowered into the can and is allowed to touch it, (c) the ball is lowered into the can as in (a), the can is touched momentarily with the finger and the ball is then withdrawn.

What conclusions can be drawn from these experiments?

Lond.

7. Explain why a sheet of metal or gauze is effective in screening a body from the influence of electrostatic forces.

8. Describe how you would charge two insulated metal spheres one positively and the other negatively.

Draw a full-scale diagram showing the lines of force between two equally and oppositely charged metal spheres if the radius of each is about 1 cm. and their centres are about 12 cm. apart.

Explain what is meant by a *line of force*, illustrating your answer by reference to a line in your diagram. *J.M.B.*

9. Define *unit electrostatic charge*.

What is the magnitude of the force between two point charges each of magnitude 20 units, placed 10 cm. apart in air? *J.M.B.*

10. Describe how you would charge a gold-leaf electroscope by induction. How would the potential of the leaves vary at each stage of the process?

Two insulated spheres A and B are placed near, but not touching each other. B is connected by a wire to the cap of a gold-leaf electroscope. Explain what happens when (i) A receives positive charge, (ii) the wire is then removed, (iii) the electroscope is then momentarily touched, (iv) the wire connection is then restored? *O.C.*

11. You bring your hand near to, but not touching, the cap of a charged electroscope; afterwards you withdraw it. State and explain what happens.

12. What do you understand by (a) the potential, (b) the capacity, of a conductor? How is each affected (if at all) by the charge on the conductor?

An insulated metal plate is connected to the cap of a gold-leaf electroscope. State and explain what happens when

- (a) an earthed metal plate is placed parallel to and about 5 cm. from the charged plate;
- (b) the earthed plate is then approached nearer the charged plate;
- (c) a thin sheet of ebonite is then placed between the plates. *Lond.*

13. Describe an experiment to show that electrostatic charge resides on the outside of a hollow charged insulated conductor.

Show how you would use this fact to compare the charges on two unequal metal spheres which have been charged when in contact and then separated. State, with reasons, which of the two spheres you would expect to have the larger charge. *Lon.*

14. Two positive electric charges, of 40 and 50 units respectively, are placed 10 cm. apart in air. Calculate the force between them. What would the force be if the charges were separated by glass (dielectric constant = 8)?

CHAPTER 34

CHEMICAL EFFECTS OF THE ELECTRIC CURRENT

If we were dealing with our subject in strict historical order, we should now have to learn about the production of the electric current, which was first obtained about the year 1799. However, we will leave that point for our next chapter. Since men knew how to obtain a current, they naturally tried all sorts of experiments with it. They soon found, for instance, that *in certain cases* substances can be decomposed by dissolving them in water and then passing an electric current through the solution. A very straightforward case is that of hydrochloric acid (Fig. 34/1). A and C are carbon electrodes in a glass jar containing the dilute acid. The current enters at the anode A and leaves by the cathode C. Hydrogen collects in the test-tube T₂, over the cathode, but very little chlorine is collected in T₁, because the gas is very soluble. The process known as electrolysis, and the vessel in which it is carried out is called an electrolytic cell or voltameter.

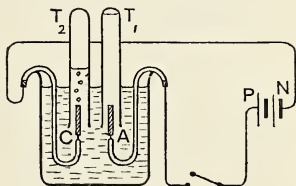


FIG. 34/1

A solution which is capable of undergoing electrolysis is known as an *electrolyte*. In some cases a substance acts as an electrolyte when in a state of fusion. Sodium chloride is an example.

Other cases previously studied—sodium chloride (in solution), copper sulphate, and ‘water’ (actually dilute sulphuric acid)—will be reconsidered later in the chapter.

Laws of Electrolysis. The subject of electrolysis was investigated by Michael Faraday. By careful measuring (in the case of gases) or weighing (solids) he found the exact quantities of the substances produced by electrolysis, and showed that these quantities were proportional to

- (i) the strength of the current and
- (ii) the time during which the current was passed.

Multiplying (i) and (ii) together we have a measure of the *quantity of electricity* that has passed through the solution and the results are summed up in Faraday's ***First Law of Electrolysis***, which states that *when a substance undergoes electrolysis, the weights of the products liberated at the electrodes are proportional to the quantity of electricity passed*.

A method of illustrating this law was described in our earlier book, and here we shall content ourselves with a brief résumé, and a sketch of the apparatus (Fig. 34/2).

The large beaker C contains very dilute sulphuric acid. Hydrogen is collected in T_2 , and it is found that the volume (and therefore the weight) is proportional to (i) the time, and (ii) the strength of the current. The latter quantity can be altered when required by adjusting the resistance R.

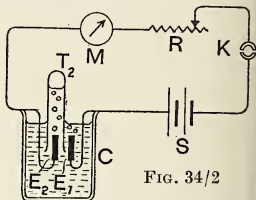


FIG. 34/2

Let us next consider the kind of experiment which led us to Faraday's Second Law. He prepared solutions of different electrolytes, and arranging them in series, passed the same current for the same time through each. Suppose, for instance (Fig. 34/3), that A is a solution of copper sulphate and B of silver nitrate, while C consists of water acidified with sulphuric acid. When the circuit is completed we find that copper is deposited on the cathode of A, and silver on that of B. At the cathode of C (as seen in the previous experiment) hydrogen is set free, and can be collected in a graduated tube. The reactions at the anodes of A and B are somewhat complicated, and for the sake of simplicity we will not consider them now. In the case of C, oxygen is given off at the anode and can be collected in another graduated tube.

At the beginning of the experiment the cathodes of A and B are weighed. The current is passed for a certain time, say half an hour, and the cathodes are weighed again. By subtraction, we find (i) the weight of copper deposited on the cathode of A and (ii) the weight of silver deposited on that of B.

The volume of hydrogen is observed, and after correcting it for temperature and pressure we can (knowing its density) calculate its weight. Similarly, we can find the weight of oxygen in the other tube. Let h, o, c and s (gm.) be the weights of hydrogen, oxygen, copper and silver respectively. Faraday found a very curious relationship between the weights, viz. $h : o : c : s = 1 : 8 : 31.8 : 108$. But why is the relationship curious?

If you look at a table of atomic weights, you will find that those of the elements concerned are respectively 1, 16, 63.6 and 108. Further, the chemical equivalent of an element is $A \div V$, where A is its atomic weight and V is its valency.¹ The valencies

of the four elements under discussion are respectively 1, 2, 2 and

Thus their chemical equivalents are—hydrogen $1 \div 1 = 1$, oxygen $16 \div 2 = 8$, copper $63.6 \div 2 = 31.8$ and silver $108 \div 1 = 108$.

Going back a few lines to the statement $h : o : c : s = 1 : 8 : 31.8 : 108$, we see that the weights of the elements liberated are proportional to their respective chemical equivalents.

Other experiments gave similar results, and so he was led to his **Second Law of Electrolysis**:—If the same quantity

Briefly, the valency of an element is the number of atoms of oxygen which one atom of the element will combine with, or displace. As one atom of oxygen combines with two atoms of hydrogen (H_2), so the valency of oxygen = 2. We may regard copper sulphate, $CuSO_4$, as being formed by the replacement of two atoms of hydrogen in sulphuric acid (H_2SO_4) by one atom of copper; so the valency of copper = 2. A comparison of the formula of silver nitrate, $AgNO_3$, with that of nitric acid, HNO_3 , shows that the valency of silver is one. The law applies not only to elements but to 'compound radicals' such as NH_4 , SO_4 , NO_3 , etc. In practice these nearly always undergo changes on liberation.

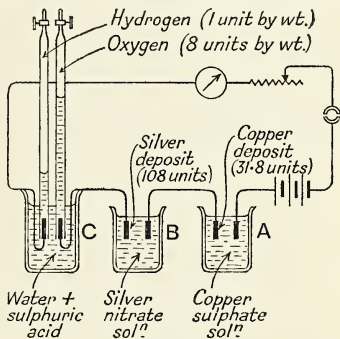


FIG. 34/3

of electricity is passed through different electrolytes, the masses of the different substances liberated at the electrodes are proportional to their chemical equivalents.

The experiment just described evidently takes a long time to carry out. It is worth noticing, however, that we can quickly obtain an example of the law from the fact that the volume of the hydrogen in C is just twice that of the oxygen. For it is known that two volumes of hydrogen combine with one volume of oxygen to form water. Thus two volumes of hydrogen are *equivalent* to one of oxygen, and so in the process of electrolysis these gases have been liberated in quantities proportional to their chemical equivalents.

Electrochemical equivalent. In what follows we shall be concerned with *quantity of electricity*, so let us be clear on the question of how this is measured.

On p. 457 we shall see that a current could be measured by its magnetic effect. In this way we obtain the 'electromagnetic unit of current,' one-tenth of which is the ampere. *The quantity of electricity conveyed by a current of one ampere in one second is known as the coulomb.*

Now, we have just seen that a quantity of electricity which would liberate 1 gm. of hydrogen would also liberate 8 gm. of oxygen, or 31.8 gm. of copper, or 108 gm. of silver, these numbers being the respective equivalent weights expressed in grams. The equivalent weight so expressed is known as the gram-equivalent. Thus the quantity of electricity needed to liberate one gram-equivalent of any substance is constant. This quantity has been found to be 96,500 coulombs.

Evidently, 1 coulomb would liberate the $1/96,500$ th part of a gram-equivalent. For copper this quantity would be $31.8 \div 96,500 = 0.000329$ gm. per coulomb, for silver $108 \div 96,500 = 0.00112$, for hydrogen $1 \div 96,500 = 0.0000105$, and so on (grams per coulomb in each case). These quantities are known as electrochemical equivalents, i.e. the electrochemical equivalent of an element is the number of grams of that element liberated by one coulomb. It would of course, have been equally correct if instead of saying 'liberated by one coulomb' we had said 'liberated by a current of one ampere flowing for one second.'

Notice that electrochemical equivalents are *proportional* to chemical equivalents. For if e_1 and e_2 are the electrochemical equivalents of two elements whose chemical equivalents are respectively E_1 and E_2 , then $\frac{e_1}{e_2} = \frac{E_1 \div 96,500}{E_2 \div 96,500} = \frac{E_1}{E_2}$.

Notice, too, that from the rate at which a metal is deposited in electrolysis, we can derive a new definition of 'ampere,' and one which will be of much more practical use than the 'electromagnetic' definition given on p. 457. Thus by careful experiment it has been found that a current of 1 ampere, passing through a suitable solution of a silver salt, causes silver to be deposited at the rate of 0.001118 gm. per sec.,¹ and so an **ampere** is now often defined as *that current which passed through a silver voltameter, will cause silver to be deposited at the rate of 0.001118 gm. per sec.*

Example 1. If the atomic weight of aluminium is 27 and its valency 3, what is its electrochemical equivalent?

Chemical equivalent = $27 \div 3 = 9$

∴ 96,500 coulombs will liberate 9 gm.

∴ 1 coulomb ,, ,, $9 \div 96,500 = 0.0000933$ gm.

∴ electro-chemical equivalent = **0.0000933 gm. per coulomb.**

A useful equation. Suppose a current of c amperes is passed for t seconds through an electrolyte, causing an element of electro-chemical equivalent z to be liberated.

Then wt. liberated by 1 ampere in 1 second = z gm.

∴ ,, ,, ,, c amps ,, 1 ,, = $z \times c$ gm.

∴ ,, ,, ,, c ,, ,, t ,, = $z \times c \times t$,,

giving this weight w , we have the equation $w = ctz$, which we shall often find useful.

Testing an ammeter. Let us see how we could use our knowledge of electro-chemical equivalents to check the readings of an ammeter.

It is from this experimental number that the number 96,500 previously given is derived. For to liberate 107.88 gm. of silver (exact weight of gram-equivalent) we need $107.88 \div 0.001118 = 96,500$ coulombs.

A rectangular strip of copper about 4 in. \times 2 in. is fastened by nails or screws to a wooden bar measuring about 4 in. \times $\frac{3}{4}$ in. \times $\frac{1}{2}$ in. as shown in Fig. 34/4b. A binding screw is inserted at S_1 . Two other strips are attached to wooden bars in the same way, and the three are suspended in a glass jar or large beaker (J in Fig. 34/4a). S_1 and S_2 , the binding screws of the outer strips, are joined by a piece of copper wire W. A circuit will finally be completed as indicated where B is a battery consisting of two 2-volt accumulators in series, K is a key, R an adjustable resistance, and A the ammeter whose readings are being tested. The central copper strip is to be a cathode, the two outer ones being connected to form a single anode.

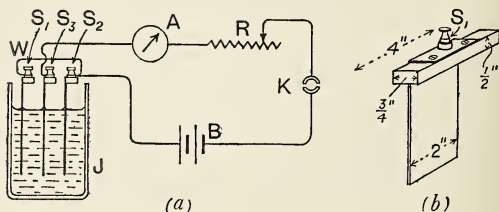


FIG. 34/4

The cathode (with its wooden attachment) is cleaned with emery paper, washed (first with water and then with methylated spirit) and dried by waving it about in the hot air over a bunsen burner. It is then carefully weighed and replaced in position in the jar. The latter is about three-parts full with a strong solution of copper sulphate to which a few drops of strong sulphuric acid have been added. The time (minutes and seconds) is noted just as the switch is closed, and the resistance is adjusted as quickly as possible so that A should read 1 amp. This reading is observed again at fairly frequent intervals, and R is adjusted if necessary to keep it to 1 ampere.

At the end of half an hour the switch is opened, and the cathode is removed, washed, dried and weighed as before.

the result is then worked out as follows:—

Wt. of cathode before expt.	.	.	.	32.348 gm.
„ „ after „	.	.	.	32.944 „
∴ wt. of copper deposited	.	.	.	0.596 „

The current as measured by the ammeter is 1 amp. Suppose *real* strength is *c* amps. It was passed for $\frac{1}{2}$ hr. = 1800 sec.

The electro-chemical equivalent of copper is 0.0003294.

∴ from $w = ctz$ we have

$$0.596 = c \times 1800 \times 0.0003294, \text{ giving } c = 1.005.$$

It would, of course, be laborious in the extreme to make a complete check of all the readings of the ammeter by the method just described. We could, however, check a few of them, including the highest and lowest, and express our results in graphical form, marking ammeter readings horizontally, and errors vertically. Our graph would then tell us what correction to apply for any given reading. We could be said to have *calibrated* the instrument.

It may be mentioned that, in the most exact work, a silver voltameter is employed. Very good results, however, can be obtained with a copper voltameter, as in the experiment just described. A source of error is the fact that unless the deposit of copper is dried very gently, it tends to oxidise.

Determination of Electro-chemical Equivalents. To find the electro-chemical equivalent of a metal such as copper, we should proceed exactly as described in the last paragraph. Here, we assumed the value of the electro-chemical equivalent, and found the strength of the current. Conversely, if we know the strength of the current (i.e. if we may assume that the ammeter is correct) we can find the value of the electro-chemical equivalent.

Example 2. In an experiment to find the electrochemical equivalent of copper, a current of 0.8 amp. was passed for 40 minutes through a copper voltameter, and caused the cathode to increase in weight from 28.731 to 29.363 gm. Find the value of the electrochemical equivalent.

$$\text{Wt. of copper deposited} = 29.363 - 28.731 = 0.632 \text{ gm.}$$

From the equation $w = ctz$, we have $w = 0.632$, $c = 0.8$ and t (in seconds) = $40 \times 60 = 2400$.

$$0.632 = 0.8 \times 2400 \times z \text{ which gives } z = 0.000329.$$

In the following example, the unknown quantity is t .

Example 3. A metal disc, 6 cm. diameter and 1 mm. thick, to be coated with a film of copper 0.01 mm. thick by electrolysis. What time will be required if a current of 2 amp. is employed? (The electro-chemical equivalent of copper is 0.000334 gm. per coulomb; the density of copper is 8.8 gm. per c.c.) Lond.*

We must first find the weight of copper to be deposited.

Area of disc (2 sides) = $2 \times \pi r^2 = 2 \times \frac{22}{7} \times 3^2 = 56.6$ sq. cm.

Area of rim = $6 \times \frac{22}{7} \times 0.1 = 1.9$ sq. cm.

Total area = $56.6 + 1.9 = 58.5$ sq. cm.

This is to be coated to a depth of 0.01 mm. or 0.001 cm.

\therefore vol. of copper to be deposited = $58.5 \times 0.001 = 0.0585$ c.c.

\therefore wt. of copper = $0.0585 \times 8.8 = 0.515$ gm.

In the equation $w = ctz$

we have $w = 0.515$, $c = 2$, $z = 0.000334$ and $t = ?$

$$\therefore t = \frac{w}{cz} = \frac{0.515}{2 \times 0.000334} = 771 \text{ sec.} = 12 \text{ min. } 51 \text{ sec.}$$

The Ionic Theory. It must not be thought that electrolysis is possible with *any* solution. If, for instance, we were to place an anode and cathode in a solution of sugar, nothing would happen—no current would flow. Electrolytes, in fact, are limited to the three classes of substances known to the chemist as *acids*, *bases* and *salts*. For simplicity, let us consider the case of the substance mentioned at the beginning of this chapter—hydrochloric acid, which is a solution of the gas hydrogen chloride.

Chemists give hydrogen chloride the *formula* HCl, meaning that the smallest particle of it that can exist—the *molecule*—consists of an atom of hydrogen, H, combined with an atom of chlorine, Cl. There is good reason to believe that when hydrogen chloride dissolves in water it *ionises*, which means that it splits up into a positively charged hydrogen atom and a negatively charged chlorine atom, thus: $\text{HCl} = \text{H}^+ + \text{Cl}^-$. These charged atoms are known as *ions*.

The charges are produced by the passage of an *electric current* (p. 402) from the hydrogen atom to the chlorine atom. You will remember that the electron is a particle of negative electricity. When the chlorine atom, previously neutral, gains

such a particle, it naturally acquires a negative charge, while the hydrogen atom, *losing* negative electricity, acquires a positive charge. Instead of $\text{HCl} = \text{H}^+ + \text{Cl}^-$, we might have written $\text{HCl} = \text{H}^{-e} + \text{Cl}^e$, e standing for an electron.

Hydrogen *ion* (H^{-e} or H^+) has very different properties from the hydrogen *atom* (simply H). Similarly chlorine ion (Cl^e or Cl^-) is very different from ordinary chlorine. For instance, it does not bleach litmus, as we can easily show by dipping litmus paper to dilute hydrochloric acid; and it does not *smell* like chlorine. In fact, it has no smell at all.

Suppose now we take some hydrochloric acid, containing hydrogen ions and chlorine ions, and into it put a positively charged plate A and a negatively charged one C—the *electrodes* of which we have so often spoken. In the absence of ions nothing would happen. There would simply be a gap in the circuit between A and C (Fig. 34/5). A would have acquired the (positive) potential of P and C the (negative) potential of N, and there the matter would rest.

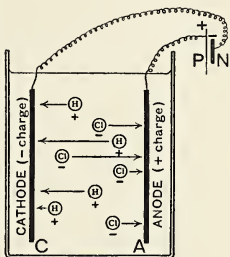


FIG. 34/5

But in the presence of ions something quite different happens. The chlorine ions, Cl^e , being negatively charged, are attracted to the positive plate (the anode). There they discharge their electrons, becoming ordinary chlorine atoms ($\text{Cl}^e = \text{Cl} + e$, $2\text{Cl}^e = \text{Cl}_2 + 2e$). We may suppose that the electrons pass through the connecting wires and battery to the cathode. What happens to them there?

Just as the negatively charged chlorine ions are attracted to the anode, the positively charged hydrogen ions (H^+ or H^{-e}) are attracted to the cathode, which is negatively charged because of the presence on it of the electrons mentioned above (those originally derived from chlorine ions). Each hydrogen ion as it reaches the cathode takes up an electron, becoming ordinary hydrogen ($\text{H}^{-e} + e = \text{H}$, or $2\text{H}^{-e} + 2e = \text{H}_2$). Thus we have a continuous flow of electrons, *inside* the liquid *from* the cathode, and *inside* the liquid *towards* the anode (where the electrons are carried by chlorine atoms) *towards* the

cathode. As the electrons represent negative electricity, the 'direction of the current' (i.e. of positive electricity) is in the opposite direction.

Some atoms (or groups of atoms) carry more than one electron. Sulphuric acid for instance, H_2SO_4 , is electrically neutral. But when it is dissolved in water, each hydrogen atom gives rise to the ions H^+ , H^+ (or H^{-e} , H^{-e}), and therefore the sulphate group must give rise to SO_4^{--} , or SO_4^{2e} , i.e. the sulphate ion must carry *two* electrons. Again, when copper sulphate, CuSO_4 , ionises, the copper ion must contain two positive charges to balance the two negative charges of the SO_4 group, i.e. copper sulphate ionises as Cu^{++} , SO_4^{--} , or Cu^{-2e} , SO_4^{2e} . We soon begin to see that *the number of electrons carried by an atom or group corresponds to its valency*.

To turn to another point, suppose we were to electrolyse a solution containing the mixed salts copper nitrate, $\text{Cu}(\text{NO}_3)_2$ and silver nitrate, AgNO_3 . Would copper be deposited at the cathode, or silver, or both?

It is found that less electrical energy is required to discharge the silver ions than the copper ions. As a result, the silver is deposited first, and later, the copper. This illustrates an important principle—that *there is a certain order of preference in which the metals are deposited*. To mention only a few of them, the order is silver, copper, hydrogen, sodium (hydrogen is not strictly speaking a metal, but resembles one in its chemical behaviour). In a similar way, there is an order of preference for the *anions*—the elements or groups which are discharged at the anode. We shall be chiefly concerned with the anions chloride, Cl^- , hydroxide, OH^- , and sulphate, SO_4^{--} , and under ordinary conditions of working we shall find they are discharged in that order—chloride most easily, then hydroxide, then sulphate.

Important cases of electrolysis. We shall now consider a few cases of electrolysis in the light of the ionic theory just discussed, beginning with *Water*. (For this experiment, the apparatus of Fig. 34/1 may be used, except that the electrodes must be of platinum, not carbon.)

If pure water is used, the action is slow in the extreme because (except to the extent of one molecule in about 500 millions) water does not undergo ionisation.

If however a little sulphuric acid is added, hydrogen and oxygen are produced at the cathode and anode respectively, in volumes of the former to one of the latter. There is no change in the amount of sulphuric acid. Let us consider carefully what has happened.

The sulphuric acid present is completely ionised ($\text{H}_2\text{SO}_4 = \text{H}^+ + \text{SO}_4^{-}$, or $\text{H}_2\text{SO}_4 = 2\text{H}^{-e} + \text{SO}_4^{2e}$). The ions thus produced travel to the electrodes, hydrogen ions to the cathode and sulphate ions to the anode.

To an extremely slight extent, water is also ionised ($\text{H}_2\text{O} \rightarrow \text{H}^+ + \text{OH}^{-}$), and the hydrogen and hydroxide ions move towards cathode and anode respectively.

At the cathode there is no complication. The only ions present are hydrogen ions, and these are discharged with the production of ordinary hydrogen. We may write $4\text{H}^{-e} + 4e \rightarrow 4\text{H} \rightarrow 2\text{H}_2$, or simply $4\text{H}^+ \rightarrow 4\text{H} \rightarrow 2\text{H}_2$.

At the anode we have numerous sulphate ions, SO_4^{-} or SO_4^{2e} , and a relatively very small number of hydroxide ions OH^{-} or OH^e . Because of the 'preferential' principle already mentioned, however, it is the hydroxide ions, not the sulphate ions, which are discharged. Corresponding to $4e$ at the cathode, we must deal with 4OH^e , and we get $4\text{OH}^e \rightarrow 4\text{OH} + 4e \rightarrow 2\text{H}_2\text{O} + \text{O}_2 + 4e$, or simply $4\text{OH}^{-} \rightarrow 2\text{H}_2\text{O} + \text{O}_2$.

Thus for every two molecules of hydrogen produced, 2H_2 , there is one molecule of oxygen O_2 , and by Avogadro's Law the volumes will be in this same ratio 2:1. The sulphate ions, with a corresponding number of hydrogen ions remain undischarged, i.e. the amount of sulphuric acid in the solution remains unchanged.

We have still to deal with one pretty obvious question. There are, relatively, so few hydroxide ions present, how can they serve to discharge sufficient electrons to maintain the current?

The answer is that as fast as hydroxide ions are discharged, they are replaced by other hydroxide ions derived from the dissociation of fresh molecules of water. Thus though the number of such ions is never large, the supply never runs out. The ions provided by the sulphuric acid which 'ferry' electrons from one electrode to the other—which carry

the current in fact—but it is those from the water which are concerned in the discharge at the anode, while at the cathode all ions are of the same sort, and no problem arises.

Copper sulphate. (i) *Using platinum electrodes.* Suitable apparatus for the experiment is shown in Fig. 34/7, a test tube T being supported over the anode by a clamp.

A gas collects in T. On examination (glowing splint test) it proves to be oxygen. The anode remains unchanged.

The cathode becomes coated with copper,¹ but no bubbles of gas are seen.

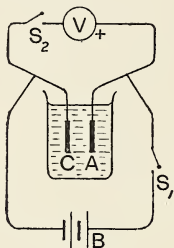


FIG. 34/6

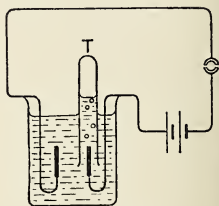


FIG. 34/7

Gradually, the blue colour of the solution becomes less intense.

Explanation. Copper sulphate and water both ionise, the latter only to a very slight extent. Thus, the cations present are Cu^{++} and H^+ , the anions being SO_4^{--} and OH^- .

At the cathode it is the copper ion which is discharged in preference to the hydrogen ion. It becomes ordinary copper which is deposited ($\text{Cu}^{++} + 2e \rightarrow \text{Cu}$).

At the anode, we have the case already discussed in connection with water + sulphuric acid. Hydroxide ion is discharged, giving rise to oxygen.

Of the ions originally present, Cu^{++} , SO_4^{--} , $(2)\text{H}^+$ and $(2)\text{OH}^-$, the first and last are steadily removed, leaving the second and third, which are the ions of sulphuric acid.

(ii) *Using copper electrodes.* If the apparatus of Fig. 3 happens to be set up, it will serve admirably. If not,

¹ Readily removed by means of nitric acid.

nplified form of Fig. 34/8 gives quite satisfactory results. In other case the anode and cathode should each be carefully cleaned, washed and dried before and after the experiment, which should proceed for about half an hour.

No gases are evolved, but a pinkish deposit of copper is seen on the cathode. On washing, and gently drying, the cathode is found to have gained weight and the anode to have lost an equal amount.¹

Explanation. The action at the cathode is the same as that already discussed. At the anode

there are three possible actions : (a) discharge of OH^- ions, (b) discharge of $\text{SO}_4^{=}$ ions, (c) conversion of copper, Cu , into copper ions, Cu^{++} , i.e. gradual solution of the anode. Of these three, it is the last which requires the least amount of energy, and which therefore takes place.

Thus, in the final result, copper is steadily removed from the

anode and added at an equal rate to the cathode, the amount of copper sulphate in solution remaining constant. For all practical purposes, the effect is as though copper 'travelled' from anode to cathode.

Practical Applications. Practical applications of electrolysis are very numerous. We have just seen, for instance, that copper may be deposited on a copper cathode, using a copper anode and a bath of copper sulphate. On a large scale this principle is used for the preparation of copper of a high degree of purity, for which there is much demand in the electrical industry.

In our earlier book (Chapter 38) we explained how electrolysis may be used to obtain an exact replica of an object, such as a medal. Gramophone records are also produced by a somewhat similar process (p. 322).

Using a solution of *silver* salt in the bath, and a silver anode, we obtain a deposit of silver. This is the basis of the

Owing to the presence of impurities, the anode may be found to have lost a little more weight than the cathode has gained.

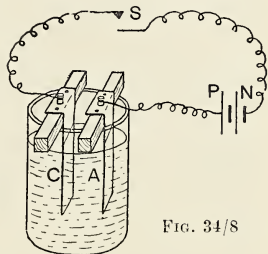


FIG. 34/8

important process of silver plating. In chromium plating (taps, car radiators, etc.) the bath contains a solution of chromic acid (from which the chromium deposit on the cathode is obtained); the anode in this case, however, does not consist of metallic chromium but of lead alloyed with antimony.

Reversal of Chemical Action. If we try to electrolyse dilute sulphuric acid, using a single Daniell cell, the results are practically nil. A few bubbles of oxygen and hydrogen collect on the anode and cathode respectively, but after a few seconds all action stops.

To understand the situation we must first consider the term *electromotive force* (E.M.F.). In a Daniell cell (or any other form of voltaic cell) a difference of potential is set up between the two plates in contact with the acid, and this difference of potential is regarded as the force moving the electricity—the electromotive force. It is measured in volts, and in the case of the Daniell cell is equal to about 1.1 volts.

Now, when we try to use a Daniell cell for the electrolysis of dilute sulphuric acid, we first get a few oxygen bubbles and a few hydrogen bubbles, as already noted. We may regard these bubbles as two *plates* of a voltaic cell, a plate of oxygen and a plate of hydrogen, and between these plates there is a difference of potential which tends to set up a current in the opposite direction to that produced by the Daniell cell. We have, in fact, what is called a *back E.M.F.* This back E.M.F. rises, and as soon as it reaches 1.1 volts the Daniell cell can no longer send a current round the circuit.

The fact that 'plates' of oxygen and hydrogen are able to set up a back E.M.F. is easily shown by the following experiment.

A battery B, consisting preferably of two accumulators in series, is connected up with a 'water voltameter' (actually dilute sulphuric acid) as shown in Fig. 34/6. The circuit may be completed or broken by means of the switch S_1 . The platinum plates, C and A, can also be connected, through a second switch S_2 , with a high resistance voltmeter V, the positive plate of the battery being connected with the positive terminal of the voltmeter.

S_2 being open, we close S_1 and allow electrolysis to proceed for a few seconds, so that oxygen collects on A and hydrogen on C. We now open S_1 and close S_2 . The voltmeter gives a reading of about 1.5 volts, but this soon falls to zero. Evidently the voltmeter has for a very short time acted as a cell, giving a current in the opposite direction to that of the current from B. As it does so, the oxygen and hydrogen bubbles recombine forming water, but the quantity of these gases is so small that the action is very quickly over.

Clearly, for electrolysis to be continued, we must use a battery of greater E.M.F. than the 'back E.M.F.' (of about 1.5 volts) set up by the oxygen and hydrogen. We could, for instance, use two Daniell cells in series (giving about 2.2 volts). But perhaps a more important point is this. Our experiment shows that it is possible,—

- (i) by means of electrolysis to produce a certain chemical change, in this case the decomposition of water into hydrogen and oxygen;
- (ii) *by reversal of the chemical change* (in this case allowing hydrogen and oxygen to recombine forming water) to produce a current.

It is true that the brief current we obtained from hydrogen and oxygen could be of no practical use, but perhaps if we used other substances we might obtain a useful continuous current. Let us repeat our experiment with one difference—for our electrodes we use strips of lead foil.

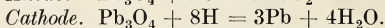
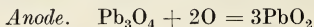
With S_1 closed and S_2 open, the anode gradually becomes black brown, owing to the formation of lead peroxide, PbO_2 . Oxygen that was previously given off at the (platinum) anode now combining with the lead. At the cathode most of the hydrogen is given off as before, but a little of it may be spent reducing (to lead) a thin film of lead oxide with which the foil is usually coated.

On opening S_1 and closing S_2 , the voltmeter records about 1.5 volts, and this does not fall so rapidly as before.¹ When the E.M.F. has fallen to zero, or near it, we can restore it by

If large pieces of foil have been used, about 8 in. \times 6 in., close together in the dilute acid, an electric bell may be put in the place of the voltmeter, and can be made to ring when S_2 is closed. The pieces of foil should be screwed to a separating strip of wood about $\frac{1}{2}$ in. wide.

opening S_2 and closing S_1 for a short time, and so on again and again. We have really made a simple *accumulator* or *storage cell*.

In a 'real' accumulator substantial lead plates are used with hollowed-out squares, slots, etc., into which red lead is compressed. On charging, this is converted into lead peroxide by the oxygen produced at the anode, and into a spongy form of lead at the cathode (where hydrogen is produced).



The lead peroxide is the positive plate of the cell, while the lead is the negative.

From this point the accumulator may be alternately discharged and charged for an indefinite number of times. We will not discuss the chemical changes that take place except to say that (i) at *discharge*, the lead peroxide and lead mentioned in the above equations are both converted into lead sulphate, PbSO_4 , some of the sulphuric acid being used up in the process. When the conversion is complete, further discharge is possible because both plates are covered with the same material; (ii) at *charge*, the lead sulphate on the anode is converted into lead peroxide, PbO_2 , while that on the cathode is reduced to lead. Thus the accumulator is once more ready for discharge.

An accumulator will have a longer life and give better service if the following points are kept in mind.

When new, the metal terminals, etc., should be smeared with vaseline, to prevent corrosion by accidental contact with the acid electrolyte. The outside should be kept dry to prevent leakage of current. Water is steadily lost by evaporation, so the accumulator should be 'topped up' every week or so, distilled water being used (ordinary water may contain metal salts in solution, which will give rise to local action).

The cell should not be allowed to give a current greater than that stated on the instructions. Otherwise, too much heat is generated, and the plates are liable to buckle, causing the paste filling to fall out. For the same reason, the cell should never be short-circuited. From time to time the specific gravity of the electrolyte should be tested with a suitable hydrometer. If it has fallen below 1.17, it is time

cumulator was re-charged. (As already mentioned, sulphuric acid is used up during discharge, causing the specific gravity to fall.) Another useful instrument is the voltmeter, which could indicate a voltage of not less than 1.85.

Questions

1. State Faraday's laws of electrolysis and explain what is meant by *electro-chemical equivalent*. Describe, with the aid of a diagram, a copper voltameter and explain how it works.

A current of 0.5 amp. is passed through a copper voltameter, and the weight of copper deposited in 15 minutes. (The electro-chemical equivalent of copper = 0.00033 gm. per coulomb.)

2. State Faraday's laws of electrolysis.

How could the correctness of the 0.5 amp. reading of an ammeter be tested by means of a copper voltameter?

If in such an experiment 0.18 gm. of copper is deposited on one of the electrodes in 20 min. what is the error of the 0.5 amp. reading? (Elec.-chem. equiv. of Cu = 0.000329 gm./coulomb).

Dur.

3. Calculate the current which, flowing for 20 minutes in a copper voltameter, deposits 0.48 gm. of copper. (Electro-chemical equivalent of copper is 0.000329 gm. per coulomb).

$$W = C t z$$

*Camb.**

4. What is meant by the *electro-chemical equivalent* of an element? Describe an experiment to determine its value for copper, and show how you would calculate the result.

A current of 2 amperes was passed for half an hour through a copper voltameter. How much copper was deposited on the cathode, if it is known that the electro-chemical equivalent of copper is 0.000329 gm. per coulomb? *Dur.*

5. Explain why an electric current flows readily through an aqueous solution of salt, but not through an aqueous solution of sugar.

6. A current is passed through three voltameters connected in series. They contain respectively (a) dilute sulphuric acid with platinum electrodes, (b) silver nitrate solution with silver electrodes, (c) sodium chloride solution with carbon electrodes. Describe the changes which take place at the six electrodes and state the relation between the quantity of oxygen liberated in one voltameter and the quantity of metal deposited in another.

J.M.B.

7. Calculate the current required to silver plate a metal tray measuring by 43 cm., on one side in 10 hours, with a coating 0.1 mm. thick. Assume the electro-chemical equivalent of silver to be 0.001118 gm. per coulomb and its density to be 10.5 gm. per c.c.

*Lond.**

8. It is required to deposit electrolytically a film of silver thickness 0.1 mm. on a spoon of surface area 20 sq. cm. If the current used is 0.1 amp., how long will the process take?

(Electro-chemical equivalent of silver 0.00112 gm. per coulomb, density of silver 10.5 gm. per c.c.) *Camb.**

9. How long would it take to give a coating of copper, 0.1 cm. thick, to a metal article of surface area 27.0 sq. cm., using a current of 0.5 amp.? Take the density of copper to be 8.8 gm. per c.c. and its electro-chemical equivalent to be 0.00033 gm. per coulomb. *C.W.B.**

10. Describe in detail how you would make a replica in metal copper of one side of a medal by an electrolytic process.

In order to deposit 15 gm. of silver on a teapot, a current 0.5 amp. was found to be suitable. For what length of time would it be necessary for the current to flow? (Electro-chemical equivalent of silver = 0.00112 gm./coulomb.) *Dur.*

11. Define *coulomb*, *ampere*, *electro-chemical equivalent*, and state the laws governing the liberation of elements by electrolysis.

A current of 2 amp. is passed for 12 min. through a water voltameter, and 280 c.c. of hydrogen measured at N.T.P. is liberated. If the chemical equivalent of copper is 31.5 and 11.2 litres of hydrogen at N.T.P. weigh 1 gm., calculate the electro-chemical equivalents of hydrogen and copper. *O.C.*

12. A current of 2 amp. is passed for 5 min. through a silver voltameter and 0.672 gm. of silver are deposited. If the chemical equivalents of silver and gold are 108 and 65 respectively, calculate the electrochemical equivalent of gold. *O.C.*

13. While testing the correctness of the 0.1 amp. reading on an ammeter, it was found that 0.4070 gm. of silver was deposited in 1 hour. Calculate the error, given that electrochemical equivalent of silver = 0.001118.

14. A current is passed for some time through two voltameters in series, the first containing copper sulphate solution, the second dilute sulphuric acid. The cathode in the copper voltameter gained 0.85 gm. in weight, while the volume of hydrogen, reduced to N.T.P., was 300 c.c. Calculate the equivalent weight of copper, assuming that 1 litre of hydrogen at N.T.P. weighs 0.09 gm.

15. A Daniell cell gives a current of 0.2 amp. for 45 min. Calculate the change in weight of both electrodes. (E.C.E. of copper = 0.00033 g. per coulomb. Equivalent weight of copper = 31.8; of zinc = 32.6.) *J.M.B.**

16. Describe a lead accumulator and explain how it works. If the polarities of the terminals of such a cell were not marked, how would you discover which was the positive pole?

CHAPTER 35

VARIOUS CELLS

THE first step towards the production of a continuous electric current was made partly as the result of a lucky accident. Luigi Galvani (1737-98), an eminent Italian scientist, was familiar with electricity generated by friction or induction, and had found that when the electricity so generated was passed through a frog's leg, it caused it to twitch. The 'lucky accident' consisted in the fact that he happened to have hung some frogs' legs on copper hooks near an iron balcony, and he noticed that whenever one of them was brought into contact with the iron, it twitched. His previous experiments made it clear that the twitching was the result of an electric discharge, but he thought (wrongly) that the electricity was stored up in the frog's leg (cf. storing in electrophorus). His fellow-countryman Volta showed, however, that the essential feature was the two different metals, copper and iron, the frog's leg merely acting as a conductor, and indicator. From this point he evolved the simple voltaic cell of Fig. 35/1, in which, however, zinc takes the place of iron. The liquid is dilute sulphuric acid. Our studies in potential (Chapter 33) suggest that if a flow of electricity takes place from C to Z when the two plates are connected, then C must be at a higher potential than Z. In this case ought we not to be able to show the difference of potential by means of a gold-leaf electroscope?

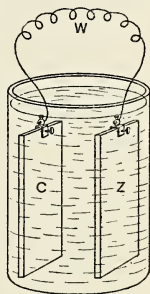


FIG. 35/1

When we make the attempt, the gold leaves show no perceptible divergence, and at first we think that perhaps the electricity is of a different *kind* from what we have met with in electrostatics. However (though we will not enter into technical details), there are ways of 'magnifying' the potential difference, and we find then that a gold-leaf electroscope is

affected in the usual way. We are, in fact, dealing with potential differences of the same kind as we met with in our work on electrostatics, but of much less magnitude.

Theory of Action of Cell. To begin with, we will consider a copper plate and a zinc plate dipping in dilute sulphuric acid with no connecting wire.

If pure zinc is used, a little goes into solution and a tiny amount of hydrogen appears at the copper plate, but the action quickly stops. Let us see why.

When zinc dissolves it becomes zinc ion ($\text{Zn} \rightarrow \text{Zn}^{-2e} + 2e$) causing the solution near the plate to be positively charged (Zn^{-2e} means that each zinc atom, originally neutral, has lost two negative units. The zinc ion is therefore positively charged.) The plate itself is negatively charged due to the presence on it of free electrons (the '+2e' of our equation). Soon it is unable to throw any further ions into solution, for being positively charged they are both held back by the negatively charged plate itself, and pushed back by the positively charged solution which surrounds it.

There are of course many hydrogen ions in the solution ($\text{H}_2\text{SO}_4 \rightarrow 2\text{H}^{-e} + \text{SO}_4^{2e}$). Some of these are driven (by the positively charged zinc-ion solution) towards the copper plate, to which they pass on their positive charges, becoming ordinary hydrogen ($2\text{H}^{-e} + 2e = \text{H}_2$). The action soon ceases, because the copper plate quickly acquires a sufficiently strong positive charge to repel any further hydrogen ions.

At this stage the copper plate, owing to its positive charge, will have acquired a certain positive potential, while the zinc plate will have a certain negative potential. The difference of potential between the two is known as the *E.M.F.* (electromotive force) of the cell, because it is this difference which, if the plates were joined by a wire, would produce movement of electricity.

So far the plates have not been connected, and the cell is said to be *on open circuit*. Now let a connecting wire be

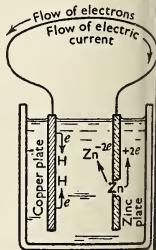


FIG. 35/2

introduced. The electrons accumulated on the zinc plate flow to the copper, neutralising the positive charge there, and making it possible for hydrogen ions to approach once more and be discharged ($2\text{H}^+ + 2e = \text{H}_2$). The student can show for himself that more zinc will now pass into solution. Thus we have a continuous *current* of electrons—from the zinc through the wire to the copper plate, where they add themselves to the hydrogen ions. In the solution itself, we may regard the flow of positive charges (carried by the hydrogen ions, H^+) from zinc to copper, as equivalent to a flow of negative charges, i.e. of electrons, from copper to zinc. Thus the electron current (negative electricity) is complete. The direction of the current of positive electricity is of course in the opposite direction.

Note that for every two electrons passed on by the zinc (an *electrical* action), one atom of zinc goes into solution (a *chemical* action). In other words, the quantity of electricity going round the circuit is proportional to the amount of zinc which goes into solution (cf. Faraday's First Law of electrolysis on p. 424). We may, in fact, say that the voltaic cell is a contrivance for converting chemical energy into electrical energy, the amount of the latter being proportional to the amount of chemical action.

Defects of the Simple Voltaic Cell. We shall presently be studying the action of the *ammeter* and *voltmeter*. For the present experiment it is sufficient to show that the former is an instrument for measuring the strength of electric current, and the latter for measuring differences of potential—much smaller differences than we meet with in our work in electrostatics.

It represents a simple voltaic cell,¹ preferably with an extra binding screw attached to the top of each plate. It is connected in series to a resistance R of about 40 ohms and an ammeter A , which should read to about 50 milliamperes. The switch S enables us to open or close the circuit at will.

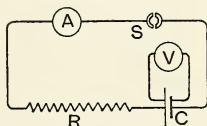


FIG. 35/3

Notice the conventional representation—long stroke and short stroke. The long stroke represents the positive plate, in this case the copper.

The voltmeter V, reading up to 2 volts, is connected across the terminals of the cell. Do not put the copper and zinc plates in the acid until all the connections have been made.

When all is ready, put the plates into the acid, but keep the switch open until the voltmeter has been read. This gives the E.M.F.—i.e. the potential difference with the cell 'on open circuit.' Now close the switch, and read the ammeter and voltmeter, repeating the readings every quarter minute. They soon show a falling off, finally becoming steady at a much lower value than originally. Now open the switch. The reading of the voltmeter slowly rises, finally reaching a value not much less than at the beginning.

How are we to account for this falling off in the strength of the current and in the E.M.F.?

Most of the trouble is due to *polarisation*—i.e. the production of bubbles of hydrogen on the copper plate. This formation of hydrogen is, of course, quite in accordance with theory, and there would be no trouble if the hydrogen *left* the copper plate at the instant it was produced. Actually, there is always a good quantity clinging to it. As a result, part of the area of the plate is put out of action, being separated from the acid by the hydrogen. It is as though the plate were steadily reduced in size, until a point is reached at which the bubbles leave the plate at the same rate as they are produced. We shall see later that a reduction in the size of the plate causes an increase in the *resistance* of the circuit, and corresponding reduction in strength of current.

Again, we have seen that when a voltaic cell is working, hydrions—(i.e. hydrogen ions) constantly reach the copper and (combining with the electrons that have come from the zinc via the wire) are turned into hydrogen. But owing to the layer of hydrogen acting as a barrier, a great many of these hydrions are unable to reach the plate. As a result, just outside the layer of hydrogen bubbles we have a layer of hydrions. These are, of course, positively charged, and they repel other hydrions which otherwise would have been able to move up to the plate. This reversal in the direction of movement of the hydrions corresponds to a tendency towards reversal in the direction of the current. It is not strong enough actually to reverse it, or even to bring it

ero, but it does greatly weaken it. We express the state of things by saying that 'a back E.M.F. is set up.'

The first remedy that comes to one's mind is to remove the copper plate from time to time and brush the bubbles away. This was actually done at one time, but *chemical* methods of removing the hydrogen were soon devised. We arrange for an *oxidising agent* to be present. As a result, the hydrogen is made to combine with oxygen and is converted to water. We shall have a good example of this almost immediately, when we come to consider the Leclanché cell.

So much for troubles at the copper plate. Now let us consider the zinc one.

According to the theory discussed on p. 32, there should be no hydrogen at the zinc plate. Actually there is a great deal, and the more impure the zinc, the greater the amount of hydrogen. How are we to account for it?

Most specimens of zinc contain considerable quantities of impurities—lead, iron and carbon, for instance. In Fig. 35/4 suppose B a particle of pure zinc and A a neighbouring particle of lead (both greatly magnified for the sake of clearness). Here, on a minute scale, we have the materials for a voltaic cell—two different metals immersed in acid,

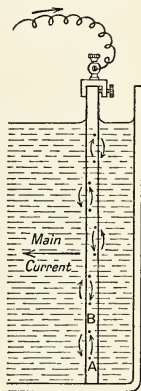


FIG. 35/4

connected not actually by a wire, but by the metal of the plate itself. The result is that we have *local action*—i.e. an electric current is set up in the direction shown by the arrows, and hydrogen appears at A. In practice we have, of course, a multitude of these local currents. Obvious results are

- (i) The effective area of the plate is reduced, because the acid cannot make contact with those parts which happen to be protected by a hydrogen bubble. Thus we have higher resistance and reduced strength of current.
- (ii) The effective area is further reduced because certain portions of the zinc are concerned in local action instead of contributing to the main current.
- (iii) Much zinc is wasted—dissolved to no useful purpose.

The most serious point of all is that this wasteful local action goes on when the cell is on open circuit—i.e. when the wire connection between the copper and zinc plates has been removed. Theoretically, when the cell is on open circuit the zinc should cease to be dissolved.

We could reduce the trouble by using much purer zinc plates, but the best remedy is found to consist in *amalgamation*. Mercury is rubbed against the zinc plate by means of a pad of cotton wool previously dipped in dilute sulphuric acid. The result is that the plate becomes covered with a thin film of *amalgam*¹ of very uniform composition. Local action takes place, of course, only when the surface of the plate is of *variable* composition.

This discussion of the reasons why the current strength of a simple cell soon declines, has of necessity been somewhat lengthy. Let us sum up briefly.

Current strength falls off because of: 1. Polarisation, 2. Local action.

1. Polarisation is caused by accumulation of hydrogen on the copper plate. This

- (i) reduces the effective area of the plate, so increasing the resistance and reducing the current;
- (ii) makes it more difficult for hydrions to make contact with the plate, and so sets up a back E.M.F. This reduces the effective E.M.F., and therefore the current, and consequently reduced.

Remedy. Removal of the hydrogen as fast as it is formed, usually by a process of oxidation.

2. Local action is caused by the fact that, owing to the presence of impurities, adjacent portions of the zinc plate differ in chemical composition and behave like separate plates. Thus numerous local currents are set up. The results are as just set out (p. 445).

Remedy. Amalgamation.

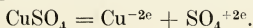
The Daniell cell. In this cell, invented by J. F. Daniell in 1836, the zinc is surrounded by dilute sulphuric acid as in Volta's simple cell. The acid, however, is contained in a porous pot, and so does not touch the copper, which in this case

¹ An amalgam is an alloy one constituent of which is mercury.

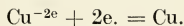
the outer vessel itself. The space between the porous pot and the copper vessel is filled with a saturated solution of copper sulphate (Fig. 35/5).

The action which takes place is as follows:—

- (i) Electrons are produced on the zinc plate and travel through the wire to the copper exactly as described in connection with the simple cell (p. 443).
- (ii) Copper sulphate in solution is ionised thus:—



- (iii) The electrons which have travelled to the copper plate convert copper *ion* into copper.



This copper is deposited on the inside of the copper vessel. (In the *simple* cell hydrogen bubbles would have been formed, producing polarisation.)

- iv) As the copper container is kept negatively charged by the constant receipt of electrons, the positively charged hydrogen ions from the inner pot move towards it, and so enter the copper sulphate solution. Thus the latter is constantly gaining hydrogen ions at the same time that it is losing copper ions—i.e. it is being gradually converted into sulphuric acid.

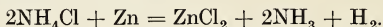
To keep the copper sulphate solution up to strength, stalks of copper sulphate are often placed on a perforated shelf just inside the copper container.

The Daniell cell has the advantage of giving a very constant E.F. Its resistance, however, is rather high, and so it does not give a very strong current (in Chapter 37 we will see that current strength varies *inversely* as resistance). Unless it is going to be wanted again very soon, it must be thrown out; otherwise the two liquids diffuse into one another through the porous pot. Thus it needs a considerable amount of attention.

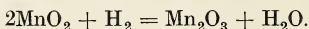
The Leclanché cell. In this cell we have zinc and carbon instead of zinc and copper, and the liquid used, instead of dilute sulphuric acid, is a strong solution of ammonium chloride. The carbon rod is contained in a porous pot, the space between rod and pot being closely packed with a mixture

of powdered charcoal and manganese dioxide. A section of the cell is shown in Fig. 35/6.

The action that takes place is often summed up briefly by saying that the ammonium chloride solution acts on the zinc according to the equation



The zinc chloride (ZnCl_2) and ammonia (NH_3) both go into solution. Hydrogen is deposited on the carbon plate where the manganese dioxide causes it to be oxidised to water



A full explanation, however, would have to take account of the part played by the electrons, as in the case of the Daniell cell just discussed.

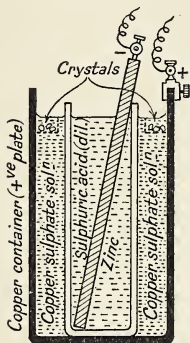


FIG. 35/5

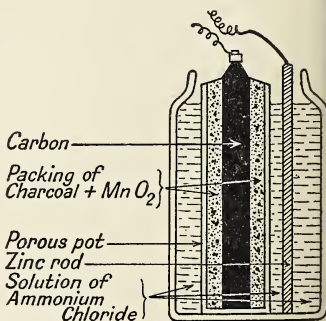


FIG. 35/6

Notice that the powdered charcoal surrounding the carbon rod may be regarded as an extension of the latter, and most of the hydrogen is produced among this powder. As the latter is in intimate contact with manganese dioxide, oxidation is greatly hastened. All the same, it is not fast enough to deal with the hydrogen at the same rate as it is produced, and so the cell is in use for some time it becomes polarised. Given a period of rest, the manganese dioxide soon oxidises the accumulated hydrogen, and the cell is then as good as ever.

From this it will be seen that the Leclanché cell is particularly suitable for *intermittent* work, such as for ringing an electric bell, or for telephone calls.

We may sum up its advantages thus:—

- (a) It is cheap and easy to maintain, seldom requiring any attention except the addition of water to replace losses by evaporation. The ammonium chloride rarely needs renewal, and when it does, costs only a few coppers. The zinc need not be amalgamated, and when the cell is on open circuit is not attacked by the ammonium chloride solution. The zinc rod lasts for a very long time, often for years.

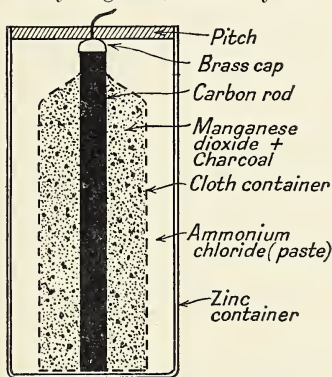


FIG. 35/7

- (b) Ammonium chloride solution is not corrosive, so no great harm is done if the cell should be upset—an important point because it is often used in houses.
- (c) Until polarisation sets in, it has a fairly high E.M.F. (about 1.5 volts).

Its chief disadvantage—unsuitability for giving a steady current for any length of time—has already been mentioned.

Dry Cell. The so-called 'dry cell' is really a modified Leclanché. It is much used for cycle lamps and other electric

torches, and quite often for ringing a house bell (instead of the ordinary Leclanché). The high-tension battery of wireless set consists of a number of dry cells.

The best way to study the construction of a dry cell is to get hold of a 'spent' one and take it to pieces. As in the Leclanché cell proper, we find a carbon rod surrounded by manganese dioxide (mixed with powdered charcoal), but these are enclosed in coarse cloth instead of in a porous pot. The zinc rod is represented by the zinc case itself, and instead of a *solution* of ammonium chloride, we have a *paste* made up of ammonium chloride, water, plaster-of-Paris and flour.

Questions

1. What is meant by the electromotive force of a cell, and what is its origin? Explain why with some cells it falls below a normal value after the cell has been in use for some time. *O.C.*

2. Explain the action of the voltaic cell in terms of the ion theory.

3. A Daniell cell gives a current of 0.25 amp. for 2 hours. How much zinc is dissolved? (The electro-chemical equivalent of hydrogen is 0.0001044 gm. per coulomb; atomic wt. of zinc 65.4.) *Brist.**

4. A Daniell cell is giving a current of 0.3 amp. In what time will 1 gm. of zinc have been dissolved? (Electro-chemical equivalent of zinc = 0.000341.)

5. Draw a sectional diagram of a Daniell cell and explain (a) why the zinc rod should be amalgamated, (b) how polarisation is prevented. *Dur.*

6. Explain the terms *local action* and *polarisation*.

Describe experiments that you could perform to illustrate the action of a primary cell, using the following apparatus: rods of copper or carbon, impure zinc and amalgamated zinc, a supply of insulated copper wire, dilute sulphuric acid, and a magnetic compass. *Camb.*

7. Describe a dry cell and give a short explanation of its mode of action.

CHAPTER 36

CURRENT MEASURED BY ITS MAGNETIC EFFECTS

FOR many years men had suspected that there was some close connection between magnetism and electricity. They had been struck by the fact that there were two kinds of electrification, just as there were two kinds of magnetic pole; and in both cases they had discovered the rule that like kinds repel and unlike kinds attract.

It was not until 1819, however, that definite evidence was obtained of the possibility of producing magnetic effects by means of an electric current. In that year the Danish professor Hans Christian Oersted was giving a lecture illustrated by experiments with a group of voltaic cells, and it chanced that part of the connecting wire (cf. W on p. 441) was in a roughly north-south direction. On the table there was a magnetic needle supported in the usual way, so as to swing in a horizontal plane. It would, of course, be pointing magnetic north and south. Oersted happened to bring the needle near the connecting wire of his battery, and noticed with astonishment that it was suddenly deflected so as to point almost east and west. On reversing the current he found that the magnet was again violently deflected, but in the opposite direction.

A year later Ampère followed up Oersted's work, and made a number of further discoveries. One of them is now always referred to as '*Ampère's Rule*,' and it may be stated thus: *Imagine a man to be swimming with the current and looking at the needle. The north pole of the latter will then be deflected towards his left hand.* Some illustrations are given in Fig. 36/1. Fig. 36/1a explains itself, and, in accordance with the rule, the north pole of the needle will be deflected towards from the paper. In (b) it is important to notice that, in order to look at the needle, the man must be swimming *his back*. Fig. (c) illustrates a case that very commonly arises in practice—a coil of wire passing both over and under the needle. We have here a combination of (a) and (b), in

each of which the north pole is deflected outwards from the paper.

Before passing on, let us notice that Ampère's rule gives an easy answer to a question which often arises—viz.: Suppose we have access to the wires connected with the terminals of a battery, but not to the battery itself. How can we tell which wire is connected to the positive terminal and which to the negative

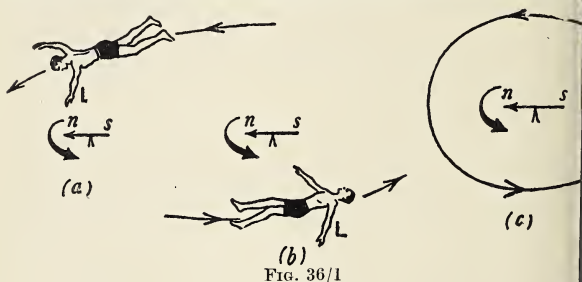


FIG. 36/1

We put a compass needle under the wire (Fig. *a*) and note which way the north pole of the needle is deflected. This gives us the position of the swimmer's left hand, and so we obtain his direction. Obviously he is swimming *from* the positive terminal.

Lines of Force. Since a magnet is deflected when near a wire carrying a current, there must be a *magnetic field* in the neighbourhood of the wire. With the help of Ampère's rule we can predict the direction of the lines of force ('the direction in which a free north pole would travel,' p. 372). Suppose the current is travelling in the direction AB (Fig. 36/2), and imagine a free north pole at N_1 . This would be driven in the direction of the swimmer's left arm—i.e. outwards from the paper. A north pole at N_2 would be driven inwards, behind the paper. If you remember that our magnetic swimmer has always to face the free north pole, we can see that the latter would follow the circular path shown on the right of the figure.

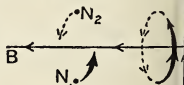


FIG. 36/2

Maxwell thought of a rule which gives exactly the same results as Ampère's, but which is much easier to apply in practice. Here it is—**Maxwell's Corkscrew Rule**. Suppose a corkscrew is being screwed along the wire in the direction of the current, then the direction in which the thumb rotates will be the positive direction of the lines of force.

Apply this to Fig. 36/2, imagining that you are making the corkscrew advance in the direction AB. You can easily see that your thumb will follow the arrows (Fig. 36/3).

Now, all this is *prediction* from Ampère's rule. Is it possible to show these lines of force by experiment?

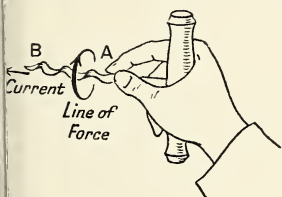


FIG. 36/3

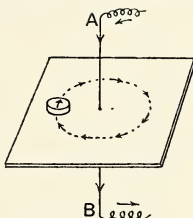


FIG. 36/4

Support a sheet of cardboard horizontally, and push a piece of rather stout copper wire, AB, vertically through it (Fig. 36/4). Connect AB to the terminals of a battery, including a resistance in circuit if necessary (you will need the help from your teacher). It is also desirable that the circuit should include a commutator switch so that the direction of the current can be readily reversed.

With the current passing in the direction AB, let a small pocket compass rest on the card and mark the direction of the needle with a pencil dot and an arrow (the latter indicating north pole). Move the compass needle into other positions on the cardboard, repeating your pencil markings in each case. The results will be found to confirm the Corkscrew Rule.

Switch over the commutator so that the current now flows in the direction BA, and put the compass successively at the old markings. The needle now points in the opposite direction.

Remove the compass and sprinkle iron filings on the cardboard. Tap gently. If the current is strong enough, the filings will arrange themselves in concentric circles (Fig. 36/5).

Coil of Wire. By means of a compass needle we can easily examine the lines of force produced by a *coil* of wire. We support a sheet of cardboard as before, and pass a coil of copper wire through it as indicated in Fig. 36/6. The lines of force are in the direction given by the Corkscrew Rule.

It is important to notice the direction of the field at the centre of the coil. If only the part A of the coil were in action, the direction of the field would be along CL—curving off to the left of the perpendicular, CP. If only the part

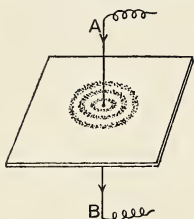


FIG. 36/5

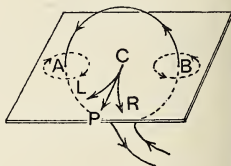


FIG. 36/6

were in action, its direction would be along CR—curving to the right. As both parts are acting, the direction is neither to the left nor right—i.e. it is at right angles to the diameter AB. In a similar way, we can prove that it is at right angles to any other diameter. To satisfy this condition, *the direction of the field, at the centre of the circle, must be at right angles to the plane of the coil*—an important result.

Notice next that the coil is behaving as a magnet would if the latter consisted of a thin circular disc of steel (of the same cross-section and thickness as the coil). For if a *field* north pole were placed at C, it would be driven along CL and that is exactly what would happen if our coil were replaced by the magnet just mentioned, provided its north pole were facing us. Such a magnet is known as a *magnetic shell*. In Fig. 36/6 the current is flowing in an anti-clockwise direction, and we are then looking at the *north* pole of the corresponding magnetic shell. To sum up, *a coil of wire*

conveying an electric current is equivalent to a magnetic shell, and if we are so placed that the current is travelling anti-clockwise, then we are facing the north pole of the shell.

This, of course, gives us a 'rule' connecting polarity with direction of current, but it is much easier to remember it in the form suggested by Fig. 31/11 on p. 367.

The Solenoid. We have just been considering the case of a single coil carrying a current. Next suppose we make our wire into a helix, or close spiral, by winding it a number of times round a cylinder such as a cardboard postal tube. The wire is supposed to be insulated. In Fig. 36/7, for simplicity, only four coils are shown.

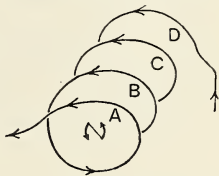


FIG. 36/7

Consider the coil A. It is equivalent to a magnetic shell, and with the current in the direction shown (anti-clockwise) the near face of this would be a north pole. The remote face is a south pole, and this would just 'cancel out' the near face of B (another north pole). Proceeding in this way, we see that the interior magnetic effects are cancelled out, and our helix behaves like a magnet of the same dimensions, with a north pole at one end (in this case the near one) and a south pole at the other.

A coil which, as a result of an electric current passing through it, is equivalent to a magnet, is known as a solenoid (from a Greek word meaning *pipe*).

From what has been said, we should expect a solenoid to be surrounded by the same sort of magnetic field as a cylindrical bar magnet. We can easily confirm this by experiment.

In a piece of cardboard we cut a rectangular hole (Fig. 36/8), of such size that our solenoid fits rather tightly into it. We then scatter iron filings on the cardboard, and tapping they arrange themselves in characteristic fashion.

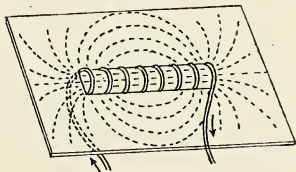


FIG. 36/8

In this case, however, as our 'magnet' is hollow, we can map the field inside. This internal field is found to be very uniform, with its direction, as we should expect, parallel to the axis of the solenoid.

The fact that a solenoid is equivalent to a magnet can also be illustrated by the 'floating battery' experiment, devised by de la Rive about a century ago. The arrangement will be sufficiently clear from the figure. There, the current appears to an observer to be travelling anti-clockwise, so the near face should be a north pole. It will therefore be repelled by a magnet held in the position shown.

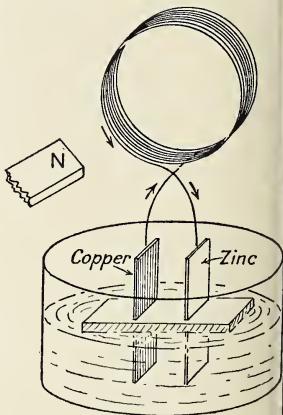


FIG. 36/9

Definition of Unit Current.

At first sight it might seem a very difficult thing to measure the strength of an electric current. But we know now that a current flowing in a coil of wire produces a magnetic field at the centre of the coil, and the stronger the current, the stronger the field. No magnetic field is something that can be measured (p. 385). It should be possible, therefore, to measure the current that produces it. It is fairly obvious that the *strength* of the magnetic field at the centre of a coil must depend on

- (i) the *strength* of the current;
- (ii) the *radius* of the coil—the bigger the radius, the weaker will be the effect at the centre;

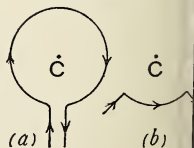


FIG. 36/10

(iii) the *length of arc* through which the current flows.

Thus in Fig. 36/10a where the current makes a complete circle, the magnetic effect at C will be greater than in Fig. 36/10b, where the current passes through only a quarter of the circle.

These ideas lead us to a definition of the **electromagnetic unit of current**, which is *that current which, flowing through 1 cm. length of an arc of radius 1 cm., produces a magnetic field of unit intensity (i.e. of one oersted) at the centre.*

The unit just defined is too large for most practical purposes. The practical unit—the familiar *ampere*—is one-tenth of it. As an exercise on the connection between current strength and magnetic field, let us work out the following:—

Example. *A current passes through three circular turns of insulated wire, side by side, of radius 10 cm. The field at the centre is 0.8 oersteds. What is the strength of the current?*

Length of wire in each circle = $\frac{2\pi}{1} \times 20$ cm., and in 3 circles

$$= 3 \times \frac{2\pi}{1} \times 20 = 188.6 \text{ cm.}$$

Suppose strength of current = x e.m. units.

Now 1 e.m. unit flowing through 1 cm. of arc of radius 1 cm. produces field of 1 oersted at centre.

$\therefore x$ e.m. units flowing through 188.6 cm. of arc of radius 1 cm. produce field of $x \times 188.6$ oersteds.

But radius is 10 cm.

\therefore field will be $\frac{x \times 188.6}{10^2}$ oersteds (inverse square law).

But field = 0.8 oersteds

$$\therefore \frac{x \times 188.6}{10^2} = 0.8$$

$$\therefore \quad x = 0.424 \text{ e.m. units} \\ = 4.24 \text{ amps.}$$

A Convenient Formula. From the definition given above we can easily derive a convenient formula.

Suppose a current of C amps. $\left(= \frac{C}{10} \text{ e.m. units} \right)$ flows through n coils of radius r cm., and that the field at the centre of the coil = F oersteds.

The total length of coiled wire through which the current passes is $2\pi r \times n = 2\pi rn$ cm. Now, 1 e.m. unit passing

through 1 cm. of arc of radius 1 cm. produces a field of oersted at the centre.

$\therefore \frac{C}{10}$ e.m. units passing through $2\pi rn$ cm. of arc of radius

1 cm. produce a field of $\frac{C}{10} \times 2\pi rn = \frac{\pi nr C}{5}$ oersteds at the centre.

But the radius is r cm., so assuming the law of inverse square the field will be $\frac{\pi nr C}{5} \times \frac{1}{r^2} = \frac{\pi n C}{5r}$ oersteds.

i.e. $F = \frac{\pi n C}{5r}$, or $C = \frac{5rF}{\pi n}$.

If we had used this formula in the preceding exercise, we should have had $F = 0.8$, $n = 3$, $r = 10$, $C = ?$

$$\begin{aligned} \therefore C &= \frac{5 \times 10 \times 0.8}{3\frac{1}{2} \times 3} \\ &= 4.24 \text{ amps. as before.} \end{aligned}$$

The Tangent Galvanometer. The tangent galvanometer is a current-measuring instrument based on the principles we have been discussing. It consists of a number of turns of insulated wire wound on a circular wooden frame. At the centre of the coil a *short* magnetic needle is suspended. In order that accurate readings may be taken, a long aluminium pointer is fixed at right angles to the needle. This pointer moves over a scale marked (usually) in degrees, and, to prevent errors due to parallax, a circular strip of mirror glass is placed on the inner side of the graduations. In taking a reading, the eye is so placed that the aluminium pointer just 'covers' its image (cf. magnetometer p. 387).

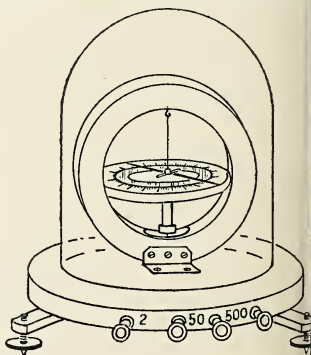


FIG. 36/11

The ends of the coil of wire are brought under the base of the instrument and soldered to the bases of two brass terminals. When we wish to measure a current, we join the ends of the wire carrying it to the brass terminals, thus including the galvanometer coil in the circuit. In practice there are usually *four* terminals, the three spaces between them being marked 2, 50 and 500. By selecting the proper pair of terminals, we can make the current pass through either 2 coils, or 50, or 500. If, for instance, we were dealing with a rather strong current, we should join up to the terminals standing left and right of '2,' so as to send it through only 2 coils.¹

The instrument is levelled by adjusting three screws which support the base, on the upper side of which a spirit-level is fitted. Before passing the current, we must arrange that the plane of the coil shall be in the magnetic meridian. We do this by gently turning the instrument until the zero of the scale is just under the end of the aluminium pointer. When we are making this adjustment it is, of course, the rule that turns—the needle continues to point north and south and the aluminium pointer east and west.)

Suppose we wish to compare the strengths of two currents (this is the purpose for which the tangent galvanometer is ordinarily used). Choosing a given pair of terminals—say those left and right of the '50'—we pass each current in turn through the instrument. In each case we read both ends of the needle, and take the average. Suppose the deflection is 44° for the first current and 22° for the second.

As we shall show presently, the required ratio is *not* $\frac{2}{1}$ but $\frac{\tan 44^\circ}{\tan 22^\circ} = \frac{0.966}{0.404} = \frac{2.39}{1}$.

If we possess certain information about the galvanometer, we can find the actual strength of each current in amperes.

It might be thought that by sending the current, say, through 250 turns, the effect would be 250 times as great as when it is sent through only 2. Actually, however, the *resistance* is very much greater in the former case, partly because of the greater length of wire, and partly because of its smaller diameter. This higher resistance means a smaller current, and so there may be little or no advantage in using a larger number of turns. There is a very marked advantage, however, if the resistance of the galvanometer is small compared with the total circuit resistance.

To get all this clear we must now consider the theory of the instrument.

Suppose NS represents the plane of the magnetic meridian in which, it will be remembered, the coil was set before the current was passed. On passing this current, a magnetic field is set up which at O, the centre, is at right angles to the plane of the coil and is equal to $\frac{\pi n C}{5r}$, where n is the number of turns, C is the strength of the current in amperes, and r is the radius of the coil (p. 458).

The needle is short, so we may take it that at P and Q the field also has this value approximately, and that the direction of the field is at right angles to NS.

Suppose the strength of the earth's field is H , the pole strength of the needle is m , and the angle through which the needle is deflected is θ .

The earth's attractive force on the pole of the needle is mH , and the *moment* of this force (p. 77) is $mH \times NP$.

The field produced by the current attracts the pole of the needle with a force of $m \times \frac{\pi n C}{5r}$, and the *moment* of this force is $\left(m \times \frac{\pi n C}{5r}\right) \times ON$.

As the needle is in equilibrium, the two moments we have mentioned must be equal.

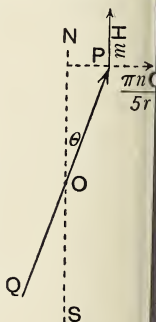


FIG. 36/12

$$\text{i.e.} \quad mH \times NP = \frac{m\pi n C}{5r} \times ON$$

$$\begin{aligned} \therefore C &= \frac{5rH}{\pi n} \times \frac{NP}{ON} \\ &= \frac{5rH}{\pi n} \tan \theta. \end{aligned}$$

Now r (radius of coil), H (horizontal component of earth's

eld), $\pi (= 3\frac{1}{2})$ and n (number of turns) are all constants, so we may write

$$C = k \tan \theta$$

when k is a constant equal to $\frac{5rH}{\pi n}$.

If we are comparing the strengths of two currents, C_1 and C_2 , then $C_1 = k \tan \theta_1$ and $C_2 = k \tan \theta_2$ (where θ_1 and θ_2 are the deflections which they respectively produce), so

$$\frac{C_1}{C_2} = \frac{\tan \theta_1}{\tan \theta_2}$$

i.e. the strengths of the currents are proportional to the tangents of the angles of deflection, as already mentioned.

If we know the values of r , H , and n , we can find the actual strength of the current. Thus suppose r (mean radius of coil) is 10 cm., $H = 0.18$ oersted and $n = 50$, then

$$k = \frac{5rH}{\pi n} = \frac{5 \times 10 \times 0.18}{3.14 \times 50} = 0.0573.$$

Thus a current which gave a deflection of 22° would have a strength ($k \tan \theta$) of $0.0573 \times \tan 22^\circ = 0.0573 \times 0.404 = 0.023$ amps.

k is known as the *reduction factor* of the instrument. It may be found in the way already indicated, but it is difficult to measure r with any great exactness. A better method is to pass through the instrument a current of known strength (either measured by another instrument, or by the electrolytic method described on p. 429), and observe the deflection. Suppose, for instance, that a current of 56 milliamps ($= 0.056$ amp.) gave a deflection of 45° . Then

$$C = k \tan \theta.$$

$$\therefore k = \frac{C}{\tan \theta} = \frac{0.056}{\tan 45} = \frac{0.056}{1} = 0.056.$$

Moving-coil Galvanometer. In the tangent galvanometer the magnet (the needle) is small and light, while the coil is relatively very large and heavy. Naturally, when a force is set up between the two, it is the needle which moves, while the coil remains stationary. If, however, we so arrange the parts that the magnet is heavy, while the coil is relatively

light, any force set up between the two will cause the coil to move, while the magnet retains its position. An instrument constructed on this principle is known as a *moving-coil galvanometer*.

The essential parts are shown in Fig. 36/13. The magnet NS is cylindrical in form (this type shows little tendency to become demagnetised). Between its poles the coil is suspended by a thin wire of phosphor bronze, the latter being connected at its upper end with a setting screw connected in turn with an arm which leads to the terminal T_1 . The other end of the coil is connected, through a spring of phosphor bronze, with a terminal T_2 . At the centre of the coil there is a core of soft iron, which serves to concentrate lines of force within the coil.

Suppose a current enters at T_1 , as shown by the arrows in the figure. This will traverse the coil in a counter-clockwise direction, and so the front face of the coil will become the equivalent of a magnetic north pole. This face will therefore move towards the south pole of the fixed magnet. As it does so, the upper wire twists, and a stage is reached at which the tendency of the coil to turn towards the south pole is balanced by the torsion force set up in the wire. It can be shown that the angle through which the coil moves is proportional to the strength of the current.

We need some means of measuring this angle. An obvious method seems to be to solder a rather long, light pointer to the wire, so that as the latter turns the pointer can move over a suitably placed horizontal scale (Fig. 36/14). This is sometimes done. A better method, however, is to make use of a much longer pointer of *light*. A small mirror, plane slightly concave, is attached to the wire (M in Fig. 36/1

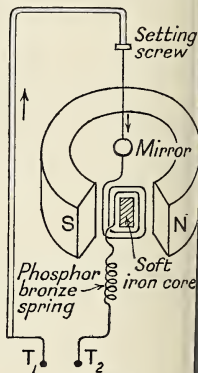


FIG. 36/13



FIG. 36/14

A lamp is placed about a metre away, and a narrow beam of light from this falls on the mirror, and is reflected from it to semi-transparent scale. A very small movement of the mirror causes a much larger movement of the spot of light on the scale.

By using a coil with a large number of turns in conjunction with a strong magnet, together with a suspending wire as thin as possible, galvanometers such as we are describing can be made to be extremely sensitive. Indeed, it is often necessary to employ a *shunt*—i.e. a conductor placed across the terminals T_1 and T_2 (Fig. 36/13), so that only a certain

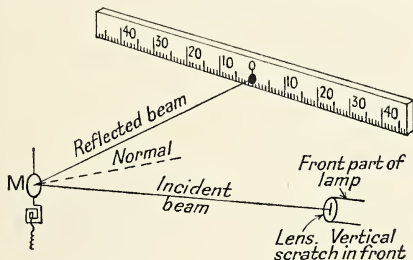


FIG. 36/15

own fraction of the current, perhaps one-thousandth, actually passes through the coils. There is usually a choice of shunts, and this fraction can then be varied according to the strength of the current. The way in which a shunt acts will be discussed more fully in the next chapter (p. 476).

Besides being much more sensitive than the tangent galvanometer, the moving-coil instrument has another great advantage. Owing to the fact that the magnetic field in which its coil rotates is so strong, other fields can be neglected in comparison. Thus (unlike the tangent galvanometer) it need not be set in any particular direction with regard to the earth's field, and its readings are not affected by the presence of neighbouring pieces of iron, or of magnetic fields due to the presence of magnets, electromotors, etc. It is (usually) 'dead-beat'—i.e. the coil comes to rest very quickly—and so a reading is soon taken.

On the other hand, the tangent galvanometer is cheaper, it needs no shunting, and it is not so easily damaged by the rather rough usage that a school instrument sometime receives. Further, one is saved the time and trouble required by the lamp-and-scale adjustments connected with the mirror galvanometer.

The Ammeter. One kind of ammeter (the word is a shortened form of ampere-meter) is a modification of the moving-coil galvanometer. The chief differences are that (i) the moving coil is pivoted on jewelled bearings instead of being suspended, and (ii) the movement of the coil is controlled by a spring, instead of by the torsion set up in the suspension wire.

Fig. 36/16 gives an idea of the general arrangement. The current enters at T_1 , but most of it is shunted to T_2 . A small fraction of it enters the instrument, and passes into the movable coil C , by means of a fine, flexible conductor. It passes out of it in a similar way, and leaves the instrument at T_2 . Inside the coil, but not touching it, is a core of soft iron to concentrate the lines of force. s is the spiral spring resisting the movement of the coil, and connected with the coil is an index finger moving over a graduated scale.

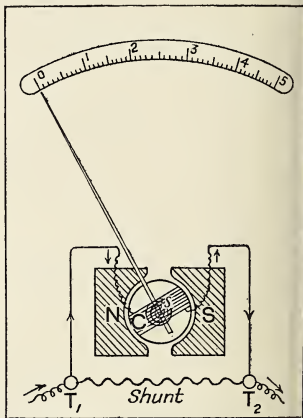


FIG. 36/16

To *calibrate* the instrument—i.e. to secure that the reading shall be correct—it is placed in series with a standard instrument,¹ and the scale is so marked that the two instruments give the same reading.

Ammeters are much used by practical electricians. The

¹ The standard instrument is itself calibrated by the electrolytic process described on p. 429.

are not so sensitive as the galvanometers previously described, but they have the great advantage of being readily portable and of being very quickly read.

Questions

A

1. A current flows as shown in the diagram. Indicate the direction of the lines of force in the neighbourhood of AB, BC, CD and DE.

2. AB and CD are two coils so suspended from stands as to be free to move towards or away from one another. Looked at from C, the current in AB is flowing clockwise, and looked at from D the current in CD is also flowing clockwise. Indicate the polarity of the two solenoids and decide whether they are attracting or repelling one another.

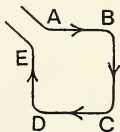


FIG. 36/17

3. State how the strength of the magnetic field at the centre of a flat circular coil depends on (a) the number of turns, (b) the radius of the coil. *J.M.B.*

4. The coil of a tangent galvanometer contains 100 turns with a mean radius of 10 cm. When a certain current is passed through it, a deflection of 45° is obtained. Find the strength of the current. ($H = 0.18$ oersted.)

B

1. Describe an experiment to demonstrate the nature of the magnetic field associated with a circular coil of wire carrying electric current. Give a diagram of the field and state a rule relating the direction of the current and the field.

2. Explain, by the aid of diagrams, the use of such a coil to measure an electric current. *C.W.B.*

3. Explain how you would show that a flat coil (such as that of a tangent galvanometer) through which a current is passing, acts as if one face is a north pole and the other a south pole. *Camb.**

4. Give a labelled diagram of a galvanometer in common use and explain the mode of action of the instrument. *Lond.**

5. In a certain tangent galvanometer a current of 2 amp. gives a deflection of 45° . What current would be required to give a deflection of 54° ? What deflection would be produced by a current of 1 amp.? ($\tan 54^\circ = 1.38$ and $\tan 26.6^\circ = 0.500$.)

5. If the coil of a tangent galvanometer has 4 turns of radius 8 cm., what is the current in amperes which will produce deflection of 45° ? ($H = 0.18$ oersted.) *Lond.**

6. How would you find the reduction factor of a tangent galvanometer experimentally?

A tangent galvanometer has three sets of windings with 5, 20, and 50 turns respectively, all of the same radius. A current of 0.1 ampere gives a deflection of 45° with the 20-turn coil. What will be the deflection observed when the same current passes through (a) the 5-turn, (b) the 50-turn winding? *Oxf.*

7. What is meant by the *reduction factor* of a tangent galvanometer, and how would you find it experimentally?

Calculate the reduction factor of a galvanometer with 10 turns of radius 10 cm., if $H = 0.18$ oersted.

8. Seven circular coils of insulated wire, side by side, have a diameter of 20 cm. When a certain current is passed through them, a field of 0.5 oersted is produced at the centre. Find the strength of the current.

9. A tangent galvanometer used in London, where $H = 0.18$, gave a deflection of 45° when a certain current was passed through it. What deflection would it give for the same current in Sydney, Australia, where $H = 0.25$? Your working must be clearly shown. ($\tan 36^\circ = 0.72$).

10. Three currents are passed in succession through the same tangent galvanometer, and are found to give deflections of 45° , 60° and 90° respectively. The first current is known to have a strength of 0.58 amp. Find the strengths of the second and third currents.

11. Describe, with a diagram, a simple type of moving-coil galvanometer.

Explain how this instrument measures currents.

On what factors does the sensitiveness of such a galvanometer depend? *Oxf.*

CHAPTER 37

OHM'S LAW AND ITS APPLICATIONS

Both in electrostatics and in current electricity we have frequently considered flow of electricity as being due to difference of potential, just as flow of water is due to difference of level. In the present chapter we shall consider the subject in greater detail, and more especially we shall see how the quantities concerned can be measured.

An obvious question to be answered if possible is this: *To what extent* does flow of electricity vary with difference of potential? If, for instance, A and B (Fig. 37/1) are points on a

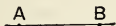


FIG. 37/1

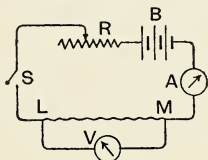


FIG. 37/2

conductor, and we have a current of 3 amps. when the potential of A is higher than that of B by 1 volt, shall we obtain a current of 6 amps. if we increase the P.D. to 2 volts?

To answer this question we set up apparatus as represented schematically in Fig. 37/2. B is a battery consisting of three 2-volt accumulators connected in series, and R is a variable resistance. LM consists of a good length, about metres, of No. 22 manganin wire. The difference of potential between its ends is measured by the voltmeter V, and the current flowing through it by the ammeter A.¹ S is a switch by which we can readily open or close the circuit.

Starting with a fairly high resistance, we open the switch a few seconds and read A and V. We now reduce R and

Strictly A records the total of the two currents that have passed through LM and V, but as V has a very high resistance (p. 475, footnote) the current passing through it is negligible.

read A and V once more. Proceeding in this way, we get results such as the following:—

Current . .	0.27	0.43	0.60	0.75	1.00	1.22
P.D. . .	1.11	1.76	2.48	3.05	4.13	5.00
$\frac{\text{Current}}{\text{P.D.}}$. .	0.243	0.244	0.242	0.246	0.242	0.24

The results show that the current is proportional to the potential difference. They are summed up in the following formal statement known as **Ohm's Law** (after the man who discovered the relationship). *If a conductor be maintained at constant temperature, the current passing through it will be proportional to the potential difference between its ends.*

If the current (in amperes) is represented by C , and the P.D. (in volts) by E , we may write $\frac{E}{C} = R$ where R is a constant.

This constant ratio is known as the *resistance* of the conductor. If $C = 1$ and $E = 1$, evidently $R = 1$, and so we get our unit of resistance, known as the *ohm*—i.e. the **ohm** is the resistance offered by a conductor through which a current of one ampere flows when a P.D. of one volt is maintained between its ends.

So far we have considered the relation between P.D. and current only for a portion of the circuit. Is the relationship true for the *whole* circuit, including the battery itself?

We can answer the question by the experiment illustrated in Fig. 37/3. A number of Daniell cells—say five—are arranged in series. The circuit includes a switch S , an ammeter A , reading to say 1 amp., and a resistance of about 10 ohms.

If e is the E.M.F. of one cell—i.e. the P.D. between its plates when on open circuit—then with the arrangement of Fig. 37/3 the E.M.F. of the whole battery is $5e$. We close the switch for a few seconds and note the current recorded by the ammeter.

We now *reverse* one of the cells. The E.M.F. of the battery as a whole is now only $3e$. Once more we read the ammeter.

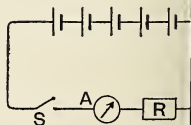


FIG. 37/3

Finally, we reverse another cell, reducing the E.M.F. of the battery to e , and read the ammeter again.

We find that, corresponding to E.M.F.'s of $5e$, $3e$ and e , the respective currents are in the ratio $5 : 3 : 1$ —i.e. current is proportional to P.D. as before. Notice the points of likeness, and of difference, between this experiment and the preceding one. In both of them we obtain different values for the current corresponding to different values of the P.D. In both, we apply our varying P.D. to something of constant resistance—to a length of wire in the one case, to the complete circuit in the other. In both the P.D. is found to bear a constant ratio to the current which it maintains.

Going back to the statement of Ohm's Law given on p. 468, we see now that by 'conductor' we must understand not simply a wire, but any part of a circuit, including battery plates and solutions; and that because the law holds good for every part of a circuit, it is true for the circuit as a whole.

Before going farther, let us familiarise ourselves with Ohm's law by working out one or two very simple examples on it. Keep in mind that, mathematically, the law may be expressed by the equation $C = E/R$, where C is the current in amperes, E the P.D. in volts, and R the resistance in ohms.

Example 1. *A piece of wire was arranged similarly to LM in Fig. 37/2. When V recorded 4.5 volts, the ammeter reading was 1.5. What is the resistance of the wire?*

Here we have $C = 1.5$ amps. and $E = 4.5$ volts. Putting these values in the equation $C = E/R$, we have $1.5 = \frac{4.5}{R}$

$$\therefore R = 3 \text{ ohms.}$$

Example 2. *In the experiment represented in Fig. 37/3, suppose each cell has an E.M.F. of 1 volt and a resistance of 8 ohms, and at the resistance R is 10 ohms, while that of A is negligible. What would be the reading of A (a) when the cells are arranged as in Fig. 37/3, (b) when one of them is reversed?*

(a) In the first case $E = 5 \times 1 = 5$ volts.

Battery ('internal') resistance $= 5 \times 8 = 40$ ohms.

External resistance $= 10$ ohms

\therefore total resistance in circuit $= 40 + 10 = 50$ ohms

$$\therefore C = \frac{E}{R} = \frac{5}{50} = 0.1 \text{ amp.}$$

(b) In this case, $E = 3 \times 1 = 3$ volts.

Total resistance in circuit is 50 ohms as before

$$\therefore C = \frac{E}{R} = \frac{3}{50} = 0.06 \text{ amps.}$$

Specific Resistance. On p. 468 we defined the *ohm*—the unit of resistance—as the resistance offered by a conductor through which a current of one ampere flows when a P.D. of one volt is maintained between its ends.

It has been found by most careful experiment and calculation that this is the resistance that would be offered by a column of mercury of certain dimensions, and the ohm is now often defined as *the resistance offered to the passage of a steady current by a uniform column of pure mercury, 106.3 cm. in length and 1 sq. mm. in cross-sectional area, at a temperature of 0° C.* This is the Board of Trade definition.

This definition evidently implies that electrical resistance depends on (i) the material, (ii) its length, (iii) its cross section and (iv) its temperature. Later in the present chapter we shall see how all these points can be proved experimentally. For the present we will merely state, without proof, the facts relating to (ii) and (iii), viz.—

- (ii) The resistance of a wire is *proportional to its length*. For instance, a piece of copper wire 100 cm. long would have twice the resistance of a 50 cm. length of the same wire.
- (iii) The resistance of a wire is *inversely proportional to the area of its cross-section*. Thus if we had two equal lengths of wire of the same material, the cross-sectional area of the first being 1 sq. mm. and of the second 2 sq. mm. the resistance of the second wire would be half that of the first.

Now it is very often necessary to calculate the resistance of a certain length of wire of given diameter. We can do this provided we know what is called the **specific resistance** or **resistivity** of the substance. This is defined as *the resistance of a conductor of that substance, 1 cm. in length and 1 sq. cm. in cross-section*.

N.B. It would be shorter, but entirely wrong, to define as 'the resistance of a cubic centimetre of that substance.' Our 'cubic centimetre' might consist of 100 yards or so of very fine wire, and its resistance would be very high.

Example 3. From the information given in the definition, p. 470, calculate the resistivity of mercury.

Length 106.3 cm. and cross-section 1 sq. mm. gives resistance of 1 ohm.

∴ length 1 cm. and cross section 1 sq. cm. (= 100 sq. mm.)

gives resistance of $\frac{1}{106.3} \times \frac{1}{100}$

$$= 0.0000941 \text{ ohms}$$

$$= 94.1 \text{ microhms}$$

i.e. resistivity = **94.1 microhms.**

N.B. A microhm = one-millionth of an ohm. Resistivity is then expressed in microhms.

The Cell as an Electricity Pump. Suppose H is a tank of water at a higher level than another tank L. If the two are connected by a pipe CDE, water flows from the higher level to the lower. In the same way, electricity passes from a point at high potential to one at a lower potential, potential in electricity corresponding to level in water.

We have had all this before, but let us now pursue the question a little farther.

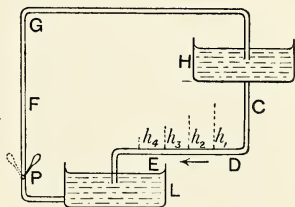


FIG. 37/4

Let us arrange a second pipe FG passing from the lower tank to the higher. Now that we have provided a complete path, is there any hope that the water will just go round and round continuously, from H to L via CDE, from L back to H via FG and so on?

Not the slightest—that would be perpetual motion. If, however, we provide a *pump* at P, the case is entirely altered. The pump acts as a source of energy, lifting the water from

the lower level to the higher, and with its help it would be quite possible to secure a continuous flow.

We have now a very helpful illustration of an electrical circuit. The electricity conveyed is represented by the water and the conducting path along which it travels is, of course, represented by the pipes. The water level has its maximum at H and its minimum at L. Correspondingly in a circuit that includes a voltaic cell, we find the highest potential at the copper plate, and this falls as we pass along the wire to the lowest value is reached at the zinc plate. In our water circuit, low-level water is converted into high-level water by the mechanical energy supplied by the pump, which corresponds in the electrical circuit to the chemical energy produced in the cell.

The resistance in the electric circuit is, of course, represented by the resistance which the pipe offers to the free flow of water. To make this obvious, let us suppose that the pipe is partly choked by growths of moss, etc., on the inside.

Such resistance will, of course, reduce the current, and the more the resistance increases, the weaker becomes the current. We are reminded very much of Ohm's Law, $C = E/R$ —the current varying inversely as the resistance.

Finally, let us consider just one portion of the pipe, say DE. The water is travelling against considerable resistance in the direction shown by the arrow. At $h_1, h_2 \dots$ we have made pinholes on the upper side of the pipe.

Thin jets of water spurt out from the holes, and we note a rapid falling off in the *height* of these jets, indicating a rapid falling off in pressure.

Example 4. A cell of e.m.f. 1.5 volts and internal resistance 3 ohms is joined to the end of a resistance of 5 ohms. What current passes through the resistance, and what is the potential difference across its ends? Lond.*

Suppose C and Z (Fig. 37/5) are the tops of the plates and C is the resistance.¹ The total resistance = $3 + 5 = 8$ ohms. Now the total fall of potential over the whole circuit is 1.5 volts and the fall is proportional to the resistance encountered.

¹ In order to bring out clearly the various points concerned, the problem has been worked out in greater detail than would ordinarily be required.

\therefore the fall in passing through the resistance CAZ is $\frac{5}{8}$ ths of the total fall—i.e. $\frac{5}{8} \times 1.5$ or 0.9375 volt (the remaining fall takes place in the cell itself and is $\frac{3}{8} \times 1.5$ or 0.5625 volt. Thus fall in external circuit + fall in internal circuit = 0.9375 + 0.5625 = 1.5 volts as we should expect.)

Now consider the resistance CAZ. The 'potential difference across its ends' we have found to be 0.9375 volts. The current passing through it is given by Ohm's Law,

$$C = \frac{E}{R} = \frac{0.9375}{5} = 0.1875 \text{ amp.}$$

We could have found the current more directly, because for the whole circuit $E = 1.5$ volts and $R = 3 + 5 = 8$ ohms

$$\therefore C = \frac{E}{R} = \frac{1.5}{8} = 0.1875 \text{ amp. as before.}$$

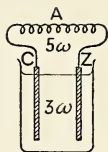


FIG. 37/5

Resistances in Series and in Parallel. The various resistances in a circuit may be arranged so that where one ends another begins. This is illustrated in Fig. 37/6. Starting from the positive pole of the battery we have—

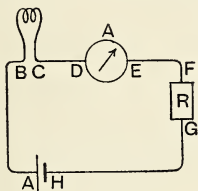


FIG. 37/6

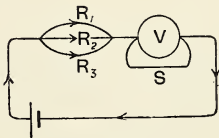


FIG. 37/7

- (i) a wire AB, (ii) a lamp BC, (iii) a wire CD, (iv) an ammeter A, (v) a wire EF, (vi) a resistance box R, (vii) a wire GH, (viii) the battery itself.

In such a case the resistances are said to be *in series*. In practice it is seldom necessary to consider the very small resistance of the wires.

Two or more resistances, however, may be so arranged as to present *alternative routes* to the current. They are then said to be arranged *in parallel*. There are two examples illustrated in Fig. 37/7.

- (i) the current divides itself between the alternative routes R_1 , R_2 , R_3 ;
 (ii) later it divides itself between the voltmeter V and the *shunt* (a sort of by-pass) S .

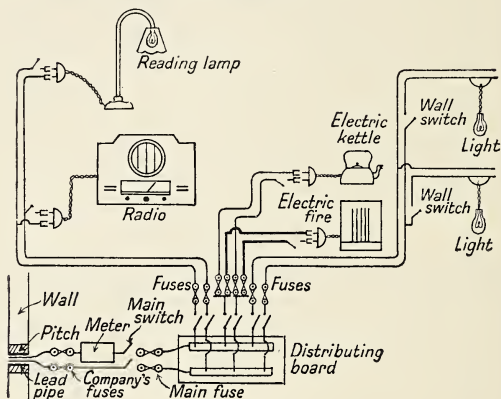


FIG. 37/8

A more elaborate example of resistances in parallel is seen in the lighting circuit of a house. In Fig. 37/8, for instance two resistances in parallel are represented by the two lights on the right. (In practice there would probably be a number of other lights on the same circuit.) On the left of the figure a reading lamp and wireless set are also seen to be connected in parallel.

When a number of resistances are in series, the *total* resistance is obtained simply by adding the separate items together. When they are in parallel, however, the matter is not quite so simple.

Suppose a current C flowing along AP reaches (at P) three resistances PFQ , PGQ and PHQ of magnitudes r_1 , r_2 , r_3 arranged in parallel. These reunite at Q , and the current then continues along the path QB (Fig. 37/9).

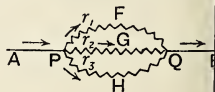


FIG. 37/9

We wish to find the single resistance that would replace r_1 , r_2 and r_3 .

Suppose the fall of potential from P to Q is E , and suppose the currents that pass through the resistances r_1 , r_2 , r_3 are respectively c_1 , c_2 and c_3 .

Applying Ohm's law to the resistance PFQ we have

$$c_1 = \frac{E}{r_1} \quad . \quad . \quad . \quad . \quad . \quad . \quad (i)$$

Similarly, considering the other two resistances we have

$$c_2 = \frac{E}{r_2} \quad . \quad . \quad . \quad . \quad . \quad . \quad (ii)^1$$

and

$$c_3 = \frac{E}{r_3} \quad . \quad . \quad . \quad . \quad . \quad . \quad (iii)$$

Adding (i), (ii) and (iii), we have

$$c_1 + c_2 + c_3 = \frac{E}{r_1} + \frac{E}{r_2} + \frac{E}{r_3}.$$

But

$$c_1 + c_2 + c_3 = C$$

and $C = \frac{E}{R}$ where R is the single resistance that would replace r_1 , r_2 , r_3 .

$$\therefore \frac{E}{R} = \frac{E}{r_1} + \frac{E}{r_2} + \frac{E}{r_3},$$

and cancelling by E , we have

$$\frac{1}{R} = \frac{1}{r_1} + \frac{1}{r_2} + \frac{1}{r_3}.$$

The result may obviously be generalised. If we have n resistances (r_1 , r_2 , r_3 . . . r_n) in parallel, then

$$\frac{1}{R} = \frac{1}{r_1} + \frac{1}{r_2} + \frac{1}{r_3} + \dots + \frac{1}{r_n}.$$

¹ Notice the important special case when there are only *two* resistances r_1 and r_2 in parallel. $c_1 = \frac{E}{r_1}$ and $c_2 = \frac{E}{r_2}$.

$\frac{c_1}{c_2} = \frac{r_2}{r_1}$, i.e. the current is shared between the two branches in *inverse* ratio of their resistances.

Example 5. Two resistances of 8 ohms and 12 ohms respectively are arranged (a) in series, (b) in parallel. Find from first principles the resistance of each arrangement. Lond.*

'From first principles' means that we must not use a formula. The first arrangement (series) would obviously give a resistance of $8 + 12 = 20$ ohms.

In the second case suppose the current C divides into two currents of c_1 (passing through the 8-ohm resistance) and c_2 (passing through the 12-ohm). Let E be the fall of potential between A and B (Fig. 37/10), and let R be the resistance of the combined arrangement.

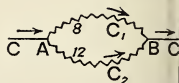


FIG. 37/10

Then $C = \frac{E}{R}$ (i)

Also $c_1 = \frac{E}{8}$ (ii) and $c_2 = \frac{E}{12}$ (iii)

But $C = c_1 + c_2$

$\therefore \frac{E}{R} = \frac{E}{8} + \frac{E}{12}$

$\therefore \frac{R}{1} = \frac{1}{8} + \frac{1}{12} = \frac{5}{24}$

giving $R = 4.8$ ohms.

N.B. Note that results such as we have just obtained can easily be verified by experiment. We should simply put the compound resistance (8 ohms and 12 ohms in parallel) in the ' r_4 ' gap in the Wheatstone bridge (p. 481) and measure its resistance in the ordinary way. It should prove to be 4.8 ohms.

Action of a Shunt. Suppose we have a moving-coil galvanometer suitable for measuring currents from 0 to $\frac{1}{10}$ amp. We wish to convert the instrument into an ammeter capable of reading, say, from 0 to 5 amps. What can we do?

Diagrammatically, the galvanometer is represented in Fig. 37/11. The current enters at the terminal T_1 and leaves at T_2 , and it must not exceed $\frac{1}{10}$ amp., or the spot of light would be turned beyond the limit of the scale.

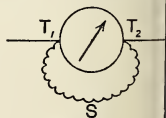


FIG. 37/11

We must so arrange matters that when the current in the

main circuit is 5 amps., only $\frac{1}{10}$ amp. actually passes through the instrument.

To do this, we connect T_1 and T_2 by a piece of wire called *shunt*, through which most of the current passes. In this case we wish to have only 0.1 amp. passing directly from T_1 to T_2 , so 4.9 amps. must pass through the shunt. Suppose the resistance of the galvanometer is R_G ohms (we shall see a little later, p. 482, how to find it). We wish to calculate that of the shunt. Suppose the latter is R_s . Now the currents passing through the two branches are inversely proportional to the resistances of these branches (footnote, p. 475).

$$\therefore \frac{0.1}{4.9} = \frac{R_s}{R_G}$$

or

$$R_s = \frac{1}{49} R_G$$

Thus the shunt must be of such dimensions that its resistance is $\frac{1}{49}$ th that of the galvanometer.

N.B. A galvanometer or ammeter, since it has to be arranged in series in a circuit, must have as low a resistance as possible.

For $C = \frac{E}{R}$, and so if our galvanometer or ammeter causes R to be greatly increased, it will cause C to be correspondingly reduced.

Let us next see how we could use our galvanometer as a *voltmeter*—i.e. as an instrument for recording the difference of potential between two points A and B in a circuit (Fig. 37/12).

Suppose we join A and B to the terminals T_1 and T_2 respectively, so that the instrument is in parallel with the main circuit. If C is the current that is shunted

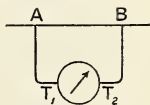


FIG. 37/12

through the instrument, then $C = \frac{E}{R}$, whence $E = CR$. Our instrument, being a galvanometer, actually records C (or can easily be made to do so). If, then, we find R once for all we have the value of $E (= CR)$.

But there is one fatal objection to all this. We have seen that the fall of potential between two points (such as A and B) in a circuit is proportional to the resistance encountered.

Evidently if we link up our galvanometer in the way just described, we materially *alter* the net resistance between A and B, because, as we have seen, the resistance of two conductors in parallel is lower than that of either one of them.

If, however, the resistance of the galvanometer 'shunt circuit' is made *very high* compared with that of the direct line AB, the net resistance is not materially altered. E.g. suppose the original resistance between A and B is 4 ohms and that of the galvanometer is 996 ohms. The resistance r of the two conductors in parallel—what we have called the 'net resistance'—is then given by $\frac{1}{r} = \frac{1}{4} + \frac{1}{996} = \frac{1000}{4 \times 996}$.
 $\therefore r = 3.98$ ohms.

So to make our galvanometer into a voltmeter, we must first put a high resistance H in series with it (Fig. 37/13). This *modified* galvanometer is then joined in parallel to the two points whose P.D. we are finding. Because of the high resistance of the alternative path, very little current is diverted.

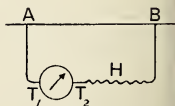


FIG. 37/13

Let us sum up.

1. To use a galvanometer as an ammeter we connect a known low resistance in parallel with the instrument so that only some known fraction of the current shall pass through it.
2. To use a galvanometer as a voltmeter we connect a high resistance in series with the instrument, so that the net resistance (and therefore the P.D.) between the two points whose P.D. is being found, shall not be materially affected.

Here are two examples.

Example 6. A galvanometer is of 5 ohms resistance, and gives a full scale deflection when the current is 15 milliams. Find the resistance needed to convert it into an ammeter reading to 15 amps. and give a diagram showing how this must be connected. J.M.B.*

We must use a shunt of such resistance (r) (Fig. 37/14) that when a current of 15 amps. passes through the complete shunted instrument, only 15 milliams. ($= 15/1000$ amps.) will pass through the

iginal coil, the difference ($15 - 0.015$) passing through the unt.

Now $\frac{\text{current through shunt}}{\text{current through coil}} = \frac{\text{resistance of coil}}{\text{resistance of shunt}}$ (footnote p. 475)

$$\therefore \frac{15 - 0.015}{0.015} = \frac{5}{r}$$

Working out, we get $r = 0.0050 \text{ ohm}$.

Example 7. A galvanometer having a resistance of 100 ohms gives a deflection of 1 scale division with a current of 0.00001 amp. Now could you adapt the instrument for use as a voltmeter with scale division for 0.01 volt? Camb.*



FIG. 37/14

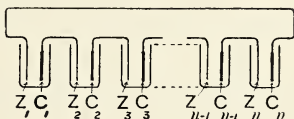


FIG. 37/15

Fig. 37/13 on p. 478, showing a high resistance H in series with galvanometer, will serve to illustrate the answer.

When $E = 0.01$, the current passing through the instrument (with H in series) must be 0.00001 amp., as this will produce a deflection of 1 scale division.

Applying Ohm's Law, $C = \frac{E}{R}$, we have $0.00001 = \frac{0.01}{R}$

$$\therefore R = 1000 \text{ (ohms)}$$

Thus the modified instrument must offer a total resistance of 1000 ohms. As the galvanometer's own resistance is 100 ohms, it will be necessary to add a resistance of 900 ohms.

Cells in Series and in Parallel. Cells are often grouped together to form a *battery*, the object being to produce a larger current. There are two principal ways in which grouping may be carried out (i) in series, (ii) in parallel. Sometimes there is an advantage in combining the two methods.

We have already referred once or twice to the 'series' arrangement. Here the positive plate of the first cell is connected to the negative of the second, the positive of the second to the negative of the third, and so on, as shown in Fig. 37/15 where Z_1, Z_2 , etc., are the negative plates and $C_1,$

C_2 , etc. the positive plates. Z_1 and C_n are then connected with the external circuit.

For simplicity let us suppose that the cells are similar and that each has an e.m.f. of E .

Let potential of $z_1 = p$

$$\begin{aligned} \therefore \quad & \text{,,} \quad \text{,,} \quad C_1 = p + E \\ \therefore \quad & \text{,,} \quad \text{,,} \quad Z_2 = p + E \text{ (because it is connected with } C_1 \text{)} \\ \therefore \quad & \text{,,} \quad \text{,,} \quad C_2 = p + 2E \text{ (because higher than that of } z_2 \text{ by } E) \\ \therefore \quad & \text{,,} \quad \text{,,} \quad Z_3 = p + 2E \text{ (because it is connected with } C_2 \text{)} \\ \therefore \quad & \text{,,} \quad \text{,,} \quad C_3 = p + 3E, \text{ and so on.} \end{aligned}$$

Reasoning in this way, we see that the potential of c_n will be $p + nE$. Thus the E.M.F. of the complete battery will be the difference of potential between c_n and z_1 (the battery being on open circuit). This difference $= (p + nE) - p = nE$, and we have the important result that *if a number of similar cells are connected in series, the E.M.F. of the battery is nE , where E is the E.M.F. of one cell and n is the number of cells.* (Whether the cells are similar or not, the E.M.F. of the battery equals the sum of the E.M.F. of the separate cells.)

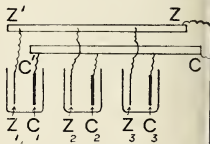


FIG. 37/16

Considering the cells as a number of resistances, the latter are, of course, in series. Thus the resistance of the battery is equal to the sum of the resistances of the various cells.

Fig. 37/16 shows (diagrammatically) a few cells connected in parallel. The positive plates are joined up by wire to form a single conductor CC' . Similarly, the negative plates are joined up to form ZZ' . The external circuit will then have its positive terminal at any point on CC' , and its negative terminal at any point on ZZ' .

Since the positive plates are all connected, they are all at the same potential; there is no 'stepping up.' Similarly, the negative plates are all at the same potential. Thus the E.M.F. of the battery is the same as that of any one of the cells.

On the other hand, all the separate positive plates form, practically, a single large positive plate, and similarly for the negative plates. The effect of this is to reduce the resistance.

As an example of the two methods of grouping, let us suppose we have four Daniell cells, each with an E.M.F. of 1 volt and an internal resistance of 8 ohms. Grouped in series, we should have a battery of E.M.F. = 4 volts, and internal resistance 32 ohms. With parallel grouping, the E.M.F. would be 1 volt and the internal resistance $8 \div 4 = 2$ ohms. Which of the two arrangements is the better depends on the *external* resistance. Let us calculate the current produced by each arrangement when the external resistance is (i) 4 ohms, (ii) 100 ohms.

(i) *External resistance small (4 ohms).*

Series grouping. E.M.F. = 4 volts. Total resistance $32 + 4$ ohms.

$$\therefore C = E/R = 4/36 = 0.11 \text{ amp.}$$

Parallel grouping. E.M.F. = 1 volt. Total resistance $2 + 4$ ohms.

$$\therefore C = E/R = 1/6 = 0.17 \text{ amp.}$$

(ii) *External resistance large (100 ohms).*

Series grouping. E.M.F. = 4 volts. Total resistance $32 + 100$ ohms.

$$\therefore C = E/R = 4/132 = 0.036 \text{ amp.}$$

Parallel grouping. E.M.F. = 1 volt. Total resistance $2 + 100$ ohms.

$$\therefore C = E/R = 1/102 = 0.001 \text{ amp.}$$

Generally speaking (as indicated by this example), when the external resistance is small compared with the internal, the parallel arrangement gives the stronger current; when it is large, the series arrangement is better.

The Wheatstone Bridge. Suppose a current of c amps. (Fig. 17) flowing along AB, divides at B into the two branches BD and BGD, which reunite at D, c_1 being the portion of the current which flows along BFD and c_2 the portion flowing along BGD. BFD is itself divided into two parts BF, FD, having resistances r_1, r_3 respectively. Now, as the potential is falling continuously along the path BFD and also along the path BGD,

there must be somewhere in BGD a point which has the same potential as F. Let G be this point, and let the resistances of BG, GD be r_2, r_4 respectively.

By Ohm's law ($C = \frac{E}{R}$ or $E = CR$) the fall of potential from B to F = $c_1 r_1$, and the fall from B to G = $c_2 r_2$.

But this fall is the same in both cases, for the potentials at F and G are equal.

$$\therefore c_1 r_1 = c_2 r_2 \quad (\text{i})$$

Again, since potential at F = potential at G, the fall from F to D = fall from G to D.

$$\text{i.e. } c_1 r_3 = c_2 r_4 \quad (\text{ii})$$

From (i) and (ii) we have

$$\frac{c_1 r_1}{c_1 r_3} = \frac{c_2 r_2}{c_2 r_4}$$

$$\therefore \frac{r_1}{r_3} = \frac{r_2}{r_4} \quad \text{or} \quad \frac{r_1}{r_2} = \frac{r_3}{r_4}$$

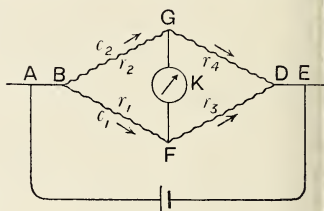


FIG. 37/17

If we know the value of r_3 and the ratio $\frac{r_1}{r_2}$, it is evident that we can easily calculate r_4 . This is the principle of the *Wheatstone Bridge*, much used for the measurement of resistance. We will now describe a common form of the instrument, making the letters correspond to the 'theoretical diagram of Fig. 37/17.

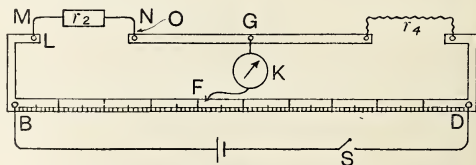


FIG. 37/18

The current enters at B, and part of it (c_1) takes the path BF, the other part going to G via the resistance r_2 . Notice that between B and G, r_2 is the only resistance that counts.

L and OG are made of stout brass or copper bar, and M and P are short pieces of fairly thick copper wire—all of negligible resistance. Similarly, r_4 is the only effective resistance between G and D.

BFD is a uniform straight wire stretched over a metre-scale, and F represents a sliding contact which can be made to touch the wire at any desired point. F and G are connected through the sensitive galvanometer K.

In discussing the theory of the bridge, it was specified that F and G must be at the same potential. We secure this by moving the sliding contact along the wire until a point is reached at which no current flows through K. In practice we begin by having a shunt attached to the galvanometer—otherwise we might get a rather damaging current passing through it. When we are in the neighbourhood of the 'null point' or 'dead point,' we remove the shunt, and we can then detect even a very faint current. When we have obtained the null point we take the reading. If BF is y , 48.2, $FD = 100 - 48.2$. Since the wire is uniform, resistance is proportional to length, so we have $\frac{r_1}{r_3} = \frac{48.2}{51.8}$.

The resistance r_2 is provided by a 'resistance box'—a piece of apparatus which can provide us with any required resistance as easily as a box of weights can supply us with any desired weight. Suppose $r_2 = 3$ ohms.

$$\text{Then} \quad \frac{r_1}{r_3} = \frac{r_2}{r_4}$$

$$\therefore \quad \frac{48.2}{51.8} = \frac{3}{r_4}$$

$$\therefore \quad r_4 = \frac{3 \times 51.8}{48.2} = 3.22 \text{ ohms.}$$

Two practical points may be noted.

1. We choose r_2 so that F is not far from the middle of the scale. A slight error in reading does not affect the ratio so seriously as it would if F were near one end.

2. After obtaining a result (as above, where $r_4 = 3.22$) exchange r_2 and r_4 , work out our answer afresh, and take the average of the two. This procedure corrects for 'end'

errors—i.e. errors arising from the fact that, owing to soldering difficulties, the ends of the wire are not *precisely* at 0 and 100.0 on the scale. It also corrects for non-uniformity of the wire. Thus our results would appear like this:—

Position	r_2	BF	FD	r_4 (calc.)	Mean
r_2 on left . . .	3.00	48.2	51.8	3.22	} 3.23
r_2 on right . . .	3.00	51.9	48.1	3.24	

Sometimes we are asked simply to find the *ratio* between the resistances of two wires. In that case the wires are put in the positions r_2 and r_4 (Fig. 37/18) and the required ratio $(length\ BF) \div (length\ FD)$. We should get a second result by interchanging them (see above), and taking the mean.

We could easily use a Wheatstone bridge to enable us to *make* a coil of given resistance. Suppose we wish to make a 2-ohm coil from a quantity of insulated wire supplied by German silver, perhaps.

First we find the resistance of a 1-metre length of the wire. Suppose this comes to 0.37 ohm. For 2 ohms it looks as though we should need $2 \div 0.37 = 5.405$ metres (540.5 cm). But any small error we made in finding the resistance of 1 metre of the wire would be increased more than 5-fold in a length of 5.405 metres, so to get a better result, cut off exactly 6 metres and measure its resistance. Suppose this comes to 2.23 ohms. By proportion, we find that the length for a resistance of 2 ohms would be $6 \times \frac{2}{2.23} = 5.381$ metres (= 538.1 cm.), and we reduce our 6-metre length to this amount.

For convenience in use, the wire would now be wound on a bobbin. Before winding, it should be doubled, as shown in Fig. 37/19, so that there shall be no magnetic effect when a current passes through it (p. 493).

One or two other uses of the Wheatstone bridge readily suggest themselves. One is that we could use



Fig. 37/19

verify the laws mentioned on p. 470 relating to the resistance of a wire. Suppose, for instance, we wish to show that the resistance of a wire is inversely proportional to the area of its cross-section. We take two wires of the same length and material, but of different diameters (measured by a micro-meter screw). Suppose they are d_1 and d_2 . Then the first has a cross-sectional area of $\frac{1}{4}\pi d_1^2$, and the second of $\frac{1}{4}\pi d_2^2$, the ratio being $d_1^2 : d_2^2$. By means of the bridge, we now find the ratio of the resistances, which should be found to be $d_2^2 : d_1^2$. Thus the law would be verified for this particular case.

Obviously, too, we could use the bridge to find the resistivity (specific resistance) of a given substance. We find the resistance of a piece of wire of that material, measure the length and diameter of the wire, and then work out the specific resistance as shown in Example 3 on p. 471.

Yet again, we could use it to show how the resistance of a wire varies with temperature. The wire in question, in the form of a rather long coil, is surrounded by a beaker of water containing a thermometer. This coil serves as, say, r_4 in Fig. 37/18, r_4 being roughly equal to it in resistance. The temperature of r_4 would remain constant, while that of r_2 could be varied as desired by heating the beaker of water, the temperature being observed on the thermometer. By means of the bridge, we could find the ratio of the resistance r_2 to r_4 for various temperatures, and as the resistance r_4 is constant, we have a measure of the change in that of r_2 . We may mention here that the increase of resistance *per ohm* in raising the temperature from 0°C. to 1°C. is called the *temperature coefficient of resistance* of the substance in question. The resistance of platinum has been observed over a wide range of temperatures. Hence a coil of platinum, enclosed in a suitable container, can be used as a *resistance thermometer*. It is used for finding a great variety of temperatures, those of furnaces, flue gases, steam superheaters, etc.

Comparison of E.M.F.s. We can find the E.M.F. of a cell by means of a voltmeter. If, however, a voltmeter is not available, we can easily *compare* the E.M.F.s of various cells by the following method, and if we know the actual value of one of the E.M.F.s we can, of course, find that of the other.

The instrument we use is called a *potentiometer*, and consists essentially of a straight uniform wire PQ of considerable resistance, stretched over a divided scale. Through the wire a steady current is flowing, say from a 4-volt accumulator B.

Suppose PQ is 100 cm. long and has a resistance of 10 ohms. If the other resistances in the circuit are negligible, we may argue that over the whole length PQ the potential falls by 4 volts, and therefore the fall per centimetre is 0.04 volt (the fall in potential is proportional to the resistance, and in the case of a uniform wire, resistance is proportional to the length). Thus at N, 25 cm. from P, the potential is lower than that of P by $25 \times 0.04 = 1$ volt.

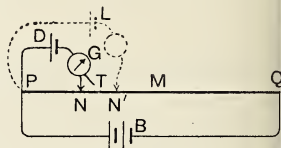


FIG. 37/20

Now suppose we connect D, the positive pole of a Daniell cell, to P. No current can flow round the Daniell circuit, for we have not yet completed the circuit by connecting up T, the other terminal. The positive pole will at once take up the potential of P.

If we take the E.M.F. of a Daniell cell to be 1 volt, the negative terminal will have a potential less by 1 volt than that of the positive terminal—i.e. 1 volt less than P. But the point N has such a potential. If, then, we make T touch the point N, there will be no current set up, because the two pieces of metal we are bringing into contact are already at the same potential. In Fig. 37/20 a galvanometer is shown as included in the circuit, and this, of course, would be unaffected.

We can sum up all this by saying that when no current flows round the Daniell circuit, the E.M.F. of the cell is equal to the P.D. between P and N (the points at which it is connected with the potentiometer wire). But, as we have seen, the P.D. between P and N is proportional to the length PN. Hence the E.M.F. of the cell is also proportional to the length.

It should now be easy to see how we make our comparison of E.M.F.s, say that of a Daniell cell with that of a Leclanché cell. With the Daniell cell arranged as shown in the figure,

move the sliding contact until the shunted galvanometer gives no deflection. We now increase the sensitiveness of the galvanometer by removing the shunt, and make any slight movement of the contact that may be necessary. Then we measure PN. Next we repeat the work with the Leclanché cell L (dotted lines in the figure) and measure PN'. Then we have

$$\frac{E.M.F. \text{ of Leclanché cell}}{E.M.F. \text{ of Daniell cell}} = \frac{PN'}{PN}.$$

To secure the best results, it is a good plan to obtain a second reading for the Daniell cell, and take the average of the two. This is to guard against the possibility that as the experiment proceeds, there may be some slight change in the D. between P and Q.

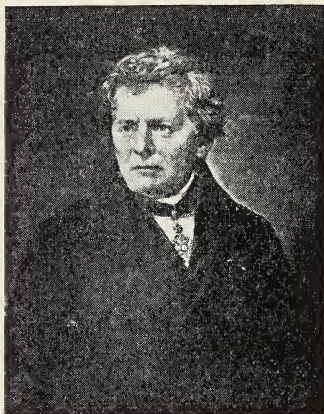
Our results might be set out thus.

Cell	Expt.	Length of PN or PN' (cm.)	Mean value	$\frac{E.M.F. (Leclanché)}{E.M.F. (Daniell)}$
Daniell	1st	26.2	} 26.25	} 1.42
"	2nd	26.3		
Leclanché	(one only)	37.5		

While the apparatus is still set up, it is worth while using to show that a Leclanché cell becomes polarised after being allowed to give a current for a little while (p. 448). After finding that for the Leclanché, the null point was at 37.5 cm., we could disconnect it from the potentiometer wire, and let it give a continuous current for a minute (say through an external resistance of 4 or 5 ohms). We could then reconnect, and find the position on the potentiometer wire once more. The process would be repeated for longer and longer 'current' periods, the fall of E.M.F. being indicated by a decreasing reading on the potentiometer wire.

§. S. Ohm. In the course of the last two chapters we have had abundant reminders that electricity is a *quantitative* science. Such things as the strength of a current, the resistance of a circuit and the electromotive force of a battery are matters of exact measurement, and Ohm's law is intimately concerned in them all. The question 'Who was Ohm?' is a very natural one.

Georg Simon Ohm was born at Erlangen in Germany in 1789. His father, a humble locksmith, had a deep respect for learning, and made great efforts to enable Simon to have a much better education than he had had himself. He even managed to keep his son at Erlangen University for an all too-short course of three terms. Simon then obtained work as a private tutor in Switzerland and saved enough money to finish his university course and pass his examinations.



G. S. OHM

By courtesy of the Director of the Science Museum, South Kensington

He had the great ambition of some day becoming University professor. For the time being he had to be content with teaching in a very ordinary school—and a very good teacher he was. Unfortunately Germany was at this time in great poverty as a result of the Napoleonic wars. Ohm's school was closed down, and he could not obtain another post. He went home, and busied himself in helping his father, now an ageing man, in the making and mending of locks. Later, the skill he acquired in this way stood him in good stead in his practical scientific work, for he was able to make much of his own apparatus.

A mathematical essay, 'written in a room without a fire,' attracted much attention, and he was offered and accepted a post as teacher of mathematics and physics at the Jesuit High School in Cologne. In his scanty spare time he carried out researches in physics, and in 1826 was able to announce his famous law.

Strange to say, it aroused much opposition among certain influential people, and the Minister of Education went so far as to state publicly that 'a physicist who professed such heresies was unworthy to teach science.'

Ohm must have felt that his dreams of a university professorship were ended now. He resigned his post, and for the next six years earned a living as well as he could, largely by taking private pupils.

Meanwhile the importance of his law was beginning to be recognised, to some extent in Germany, but to a far greater extent abroad. Sir Charles Wheatstone, for instance (a famous English physicist after whom the famous 'bridge' is named), had the highest opinion of his work. In 1833 he received an appointment at the Nuremburg Polytechnic School, and in 1841 the Royal Society awarded him the famous Copley medal in recognition of his great services to science. In 1849 came perhaps the greatest moment of his life, when, after all his struggles against poverty and undeserved censure, he was appointed a professor of the University of Munich, where he died a few years later.

Questions

1. The terminals of a cell, of E.M.F. 1.07 volts and internal resistance 0.5 ohm, are joined by 1 m. of wire of resistance 0.0025 ohm per cm. Calculate the current. What would the latter be if the length of wire were increased to 2 m.?
2. A filament lamp of resistance 4 ohms is run on a battery of E.M.F. 6 volts and an internal resistance 2 ohms. Assuming there are no other resistances in the circuit, find (a) the current strength, (b) the potential drop across the filament of the lamp.
O.C.
3. Define *ampere*, *volt* and *ohm*.
What length of wire of diameter 2 mm. and resistivity (specific resistance) 0.000049 ohm-cm. is necessary to make a resistance of 5 ohms? Describe briefly how the value can be roughly checked by experiment. Lond.

4. State *Ohm's law*, and define *specific resistance (resistivity)*. Describe how you would measure accurately an unknown resistance of about 1 ohm.

The resistance of 100 metres of copper wire of 0.50 mm diameter is 8.2 ohms. Calculate the specific resistance of copper
Camb.

5. Find the resistance of 2 metres of eureka wire, the specific resistance of eureka being 0.000044 ohm-cm., and the diameter of the wire 0.122 cm. How would you measure the resistance of this wire experimentally? *C.W.B.**

6. When a galvanometer is connected (a) in series, (b) in parallel, with a resistance of 3 ohms, and the combination is joined to a battery of constant E.M.F. and negligible resistance the currents indicated are in the ratio of 3 to 4. Calculate the resistance of the galvanometer. *O.C.**

7. A current of 2 amp. passes through a resistance consisting of two coils of resistance 4 ohms and 6 ohms respectively, arranged in parallel. Calculate (a) the equivalent resistance of the coils (b) the current in each coil. *Lond.**

8. A battery of twelve accumulators is to be charged from 230 volt D.C. mains. If the E.M.F. of each cell during charging is 2.5 volt and the internal resistance negligible, what resistance must be included in the circuit to give a charging rate of 8 amps. If a number of resistances of 100 ohm are available, how many would you use, and how would you arrange them so as to give this charging rate? Draw a circuit diagram. Give reasons why this method of charging is wasteful. *Camb.*

9. Describe in detail how you would measure an electric current by means of the tangent galvanometer.

A tangent galvanometer of resistance 1 ohm is connected in series with a resistance of 10 ohms and a battery, and a deflection of 60° is obtained. When the external resistance is increased to 38 ohms the deflection is halved. What is the internal resistance of the battery? *Camb.*

10. A cell is connected in series with a resistance of 1000 ohms and a tangent galvanometer of resistance 1000 ohms. The deflection is 45° . If a resistance of 1000 ohms be placed in parallel with the galvanometer what will the galvanometer deflection be? *Dur.**

11. State *Ohm's law*.

Nine cells in series, of total resistance 3 ohms, are connected to two points A and B, between which are resistances of 10 and 15 ohms in parallel. An ammeter shows the total current to be 2 amp. What is (a) the E.M.F. of each cell, (b) the potential difference between the points A and B, (c) the current in each resistance between A and B?

12. What is meant by *resistivity* or *specific resistance*, and how would you measure it in the case of nichrome, given a coil of wire of that material and any required apparatus?

A tungsten filament lamp dissipates 40 watts when used on 40 volt D.C. mains. What is the resistance of the lamp when in use?

Find also the length of the filament if its diameter is 0.025 mm. and the resistivity of the hot filament is 45×10^{-6} ohm cm. *Dur.*

13. State Ohm's law and define *electrical resistance*.

The terminals of a cell of E.M.F. 1.5 volts and internal resistance 0.5 ohm are connected in series with two resistances A and B. A consists of a 2-ohm coil and B consists of two resistances of 2 ohms and 4 ohms arranged in parallel. Draw a diagram of the circuit. Calculate (a) the total resistance of the circuit, (b) the current flowing through the cell, (c) the currents flowing through the separate branches of B. *Camb.*

14. A battery of E.M.F. 12 volts and internal resistance 2 ohms is connected in series with a 5 ohm resistance and a combination of resistance composed of a 1 ohm coil and a 4 ohm coil in parallel. Calculate the currents flowing through the 1 ohm and 4 ohm coils, and the voltage between the common terminals. *Dur.**

15. Two similar cells, each of E.M.F. 2 volts, are connected in series to form a battery which sends a current through an external circuit of resistance 6 ohms. The cells are then arranged in parallel, and when this battery is used the current through the same external circuit is $14/25$ of that previously. Calculate the internal resistance of each cell. *O.C.**

16. State Ohm's Law, and describe an experiment to verify

Three wires, A, B and C, each have a resistance of 12 ohms. Find the values of the resistances made by (a) connecting all three in series, (b) connecting all three in parallel, (c) joining A and B in series and connecting this composite resistance in parallel with C. *Oxf.*

17. State *Ohm's law*.

Why is the potential difference between the terminals of a cell always the same as the electromotive force of the cell?

Resistances of 2 and 3 ohms, joined in parallel, are connected to the terminals of a cell of E.M.F. 1.5 volt and internal resistance 0.5 ohm. Calculate (a) the current through the cell, (b) the difference of potential between the terminals of the cell, (c) the current in the 2 ohms coil. *Camb.*

18. A galvanometer of resistance 63 ohms is connected in circuit with a battery, the E.M.F. of which is 10 volts. The resistance of the battery and wires is 2 ohms. The galvanometer is shunted so that one-tenth of the total current passes through

Calculate (a) the resistance of the shunt, (b) the current through the battery. *O.C.**

19. A battery of internal resistance 1 ohm is connected in series with a resistance box, and galvanometer of resistance 50 ohms. The latter is shunted by a wire of resistance X ohms. When the resistance of the box is 10 ohms, the current recorded by the galvanometer is half that shown when the resistance of the box is 3 ohms. Calculate X . *O.C.**

20. A galvanometer having a resistance of 99 ohms is suitable for measuring currents from 0 to $1/10$ amp. Find the resistance of the shunt that will make the instrument suitable for measuring currents from 0 to 1 amp. *O.C.**

21. Describe an instrument for the measurement of electric current.

A galvanometer of resistance 2 ohms is graduated to read up to 0.1 amp. Explain how you would modify it into a voltmeter to read up to 10 volts. *O.C.*

22. Describe a moving-coil galvanometer, and in your diagram show the position of the coil when no current is passing.

If you were given a 2000-ohm coil and a moving-coil galvanometer, how would you compare the E.M.F. of a Daniell cell with that of a Leclanché cell? What purpose is served by the coil of high resistance?

23. Draw diagrams to show how three similar cells, each of 1.4 volts e.m.f. and 5 ohms internal resistance, should be connected to give (a) the highest possible e.m.f., (b) the lowest possible internal resistance. State both the e.m.f. and the resistance of the arrangement in each case. *J.M.B.**

24. Six cells, each of internal resistance 0.5 ohm and 1.1 volt E.M.F., are connected (a) all in series, (b) all in parallel, (c) in two parallel sets of three cells each. Calculate the current sent in each case through a wire resistance of 1.2 ohm. *Dur.*

25. State *Ohm's law*, and describe how you would verify it experimentally.

Two cells, each of E.M.F. 1.5 volts and internal resistance 2 ohms, are connected in parallel so as to supply current to a wire of resistance 4 ohms. Find the value of the current and the potential difference between the ends of the wire. *Oxf.*

26. Describe an instrument which can be used for measuring current by means of its magnetic effect.

Two accumulators, each of E.M.F. 2 volts and the same internal resistance, are connected in parallel and used to send a current through a resistance of 7.5 ohms and an ammeter of resistance 2 ohms connected in series. The current found is 0.2 amp. Calculate the resistance of each cell. *O.C.*

27. Describe a method of comparing the resistances of two wires, giving a diagram of the circuit used.

Two wires, one of copper and one of iron, each 100 cm. long and 0.070 sq. mm. in cross-section, are connected in series, and a potential difference of 2 volts is applied to the arrangement. Calculate (a) the resistance of each wire, (b) the fall of potential along the copper wire.

(Specific resistance of copper = 1.77×10^{-6} , and of iron = 10.5×10^{-6} ohms per cm. cube.) *J.M.B.*

28. Describe how you would make a 3-ohm coil if you were provided with a standard 1-ohm resistance, a reel of insulated resistance wire and any other apparatus which you might require.

If the length of the wire used in making the coil was found to be 1 metre and its diameter was 0.3 mm., what length of the same kind of wire, but of diameter 0.4 mm., would you cut off to make 3-ohm coil? *Dur.*

29. Explain how you would compare the E.M.F. of two cells by means of a potentiometer.

A cell is connected across a resistance of 1 ohm, and the potential difference across the resistance is found to be 1.0 volt. When the resistance is replaced by one of 7 ohms, the potential difference is 1.4 volts. Calculate the E.M.F. and internal resistance of the cell. *Camb.*

30. What do you understand by (a) *electromotive force* of a cell, (b) *potential difference* between two points?

You are provided with the usual potentiometer equipment, an accumulator, and a standard cell of given E.M.F. How would you use them to determine the E.M.F. of a Daniell cell?

If a known low resistance is connected in parallel with the Daniell cell, what effect will it have on the potentiometer readings? What can you deduce from this observation? *Camb.*

31. Describe how you would compare the e.m.f.s of a flashlamp battery (about 3 volts) and a Leclanché cell (about $1\frac{1}{2}$ volts) by means of a potentiometer. Give the theory of the method employed, and point out any details of experimental procedure which ensure accuracy. *J.M.B.*

CHAPTER 38

ELECTROMAGNETIC INDUCTION

UNTIL 1831 some form of voltaic cell was the only known means of obtaining an electric current. No doubt certain practical applications would have been possible, but they would have been comparatively few, and only on a small scale. On a basis of zinc, copper and sulphuric acid we could never have had supplies of electrical energy sufficient to drive our trains and tram-cars, to light our towns and heat our houses on the scale that we see round us to-day—the cost would have been far too great.

What made the new developments possible was Faraday's

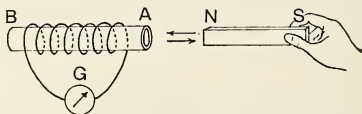


FIG. 38/1

discovery of how to convert mechanical energy into electricity and it is this conversion which we must now discuss.

We considered some of his experiments in our earlier book. In one of them (Fig. 38/1) he had a coil of wire, the two ends of which were connected with a galvanometer G. When a magnet NS was *held* in or near the coil, nothing happened. When, however, it was *suddenly moved* towards the coil, or into it, the galvanometer showed that a current was produced, and when the magnet was quickly withdrawn, a current was again produced, but in the opposite direction. The current is said to be *induced* in the coil, and the phenomenon is known as *electro-magnetic induction*.

Carrying out the experiment to-day, we should make, say, 250 turns of insulated copper wire round a cardboard tube about 4 in. long and an inch in diameter, and use a sensitive galvanometer. We could then easily show—

- (i) that the strength of the induced e.m.f. depends on the speed of movement.

(What we should observe—with the help of the galvanometer—is that the strength of the induced *current* depends on the speed of movement. But since $C = E/R$, and R is constant for the same coil, a greater or less value of C indicates a greater or less value of E —i.e. of the induced e.m.f. It is better to speak of the induced e.m.f. rather than of the induced current, because the strength of the latter depends not only on the movement of which we are speaking, but on the resistance of the circuit. Besides, an e.m.f. is induced even if the coil is open, while a current flows only if it is closed.)

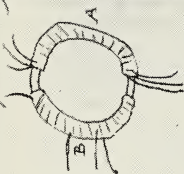
- (ii) that by using a stronger magnet, we obtain a stronger induced e.m.f.—once again, indicated by a stronger current (the easiest way to obtain a stronger magnet is to tie another magnet to the original one, with like poles together);
- (iii) that (as we should expect) it makes no difference whether the magnet is pushed into the coil, or the coil is pushed over the magnet.
- (iv) that the induced current is in opposite directions according as the magnet is being pushed into the coil, or withdrawn from it.

We must now look more closely into this question of the *direction* of the induced current. Let us first see whether we can predict it.

An electric current represents so much energy, and therefore when we produced one by moving a magnet, we must have done work—we must have met with *resistance*—when we moved NS towards the coil and also when we pulled it away (Fig. 38/1).

If we met with resistance as we moved NS towards the coil, the end A must have been acting as a north pole ('like poles repel'), and therefore, looked at from N, the current must have been anti-clockwise (Fig. 31/11, p. 367). But we must also have met with resistance in *withdrawing* the magnet. In that case the end A must have been attracting and so must have acted as a south pole. The current (looked at from N) must therefore have been clockwise.

would be from the other with all the ends of the ring
 A. on the other side but separated by an
 insulated wire wound twice in two pieces
 together and run by to about 60 feet in
 by the the distance begins with the former
 into this side with B.



Charged a battery of 16 ft. plates twice again. Made
 the end on B side one end and connected its extremity by
 a copper wire passing to a distance, and put over a magnetic
 needle (3 ft. from wire) then connected the end of one of the
 pieces on A side with battery, immediately a visible effect on needle
 & magnet of either it had an inverted position. On breaking
 connection of A side with Battery gave a disturbance
 of the needle

Our galvanometer indicates the direction of the current passing through it (needle moving to the right means that the current is entering at the ' + ' terminal). Thus we can put our conclusions to the test of experiment, and we find they are confirmed. They are summed up in what is known as **Lenz's law**, which says that *If a conductor is moved in relation to a magnetic field, the direction of the current induced in it will be such as to oppose the motion.*

A striking experiment to illustrate Lenz's law is carried out by placing a heavy copper ring on the pole of an electro-magnet (Fig. 38/2). When the current is switched on, the ring is thrown violently upwards. The switching on is equivalent to bringing a magnet near the coil, and a current is induced in the copper ring in a direction tending to push the magnet away. As the magnet is a fixture, the copper ring is pushed away instead.



FIG. 38/2

Induced Current and Lines of Force. When we look at a magnet most of us see—just a magnet. But Faraday saw a great deal more. He saw the lines of force all round it; and he came to the conclusion that the e.m.f. giving rise to the induced current depended on the *rate* at which these lines of force were being cut by the coil. This would account for the fact that the more rapidly the magnet (or the coil) was moved, the stronger was the current. It would also account for the fact that the strength of the induced current depended on the length of the magnet, because a stronger magnet would produce a greater concentration of lines of force than a weak one. All this could be true even if our coil consisted only of a single turn, and since the effect is experienced by each turn, the total effect will obviously depend on the number of turns.

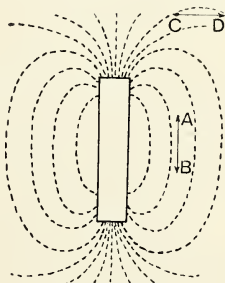


FIG. 38/3

It is quite possible to move the

coil so that it travels *in the direction of* the lines of force instead of cutting *across* them—following a path such as AB or CD in Fig. 38/3. In that case we obtain no induced current.

Coil instead of magnet. On p. 455 we saw that a coil in which a current is flowing—a *solenoid*—is equivalent to a magnet. We should therefore expect that if the magnet NS of Fig. 38/1 were replaced by a solenoid (CD in Fig. 38/4)

we should still have currents induced in AB; and this is found to be the case. Looked at from the left, the current in the solenoid would have to be anti-

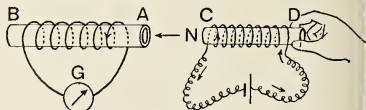


FIG. 38/4

clockwise for the end C to be equivalent to a north pole. Lenz's law tells us that as CD is pushed near to AB, the equivalent of a north pole would then be developed at A—i.e. looked at from the right the current in A would be anti-clockwise or looked at from the left, clockwise. Thus the two currents are in opposite directions.

We can arrive at the same result in another way. In Fig. 38/4a, for simplicity, only the end coils of C and A are shown.

On the 'C' coil, we have marked not only the direction of the current (at K), but (using the Corkscrew rule), the direction of the lines of force (FF). Now suppose C to be moved rapidly towards A. The induced current would have to be

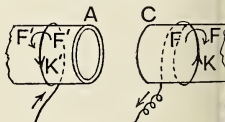


FIG. 38/4a

such as to set up an opposing magnetic field. The lines of force would therefore have to run in the reverse direction (F'F'), and the Corkscrew rule tells us that the current would be as indicated at K', i.e. in the opposite direction to that of the original current.

Turning again to Fig. 38/4, let us suppose that the 'CD' circuit includes a switch, so that a current can be started or stopped at will. The switch is open—no current. Now instead of moving CD towards AB as we did before, let

lose the switch. A current suddenly starts up, with the usual attendant lines of magnetic force.

At first sight it looks rather difficult to apply Lenz's law, but if we have not exactly moved a conductor in relation to a magnetic field, we have moved a magnetic field in relation to a conductor, which is really the same thing. Fig. 38/4a will now serve us once more. It shows the direction of the current (at K), and of the lines of force (at FF') in the 'CD' coil. An opposing magnetic field will be set up in the 'AB' coil (direction shown at F'F'), and the current to produce such a field would have the direction shown at K', i.e. it would be opposed to the original direction.

We have dealt only with the cases where CD is suddenly brought near to AB, or where a current is suddenly started in CD. You can work out the other cases for yourself (i.e. CD suddenly drawn *away* from AB, or a current suddenly *stopped* in CD). In both cases you will find that the induced current is in the same direction as the inducing current.

An important case arises when one coil CD is inside another AB (Fig. 38/5). If the switch is closed for a moment, a current, with its attendant magnetic field, is at once set up in CD. The current induced in AB will be such as to produce an opposing magnetic field, i.e. the induced current will have a direction opposite to that of the inducing current. All the results we have worked out can easily be confirmed by experiment.

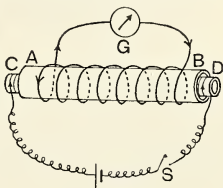


FIG. 38/5

It would be laborious, of course, if we had to *work out*, in every case, the direction of the induced current. It is well to work them out in the first instance, however, for we thereby gain a much firmer grip of Lenz's law, and we then take care to remember that

At *make*, induced current has *opposite* direction.

„ *break* „ „ „ *same* „

We have already seen that the e.m.f. of the induced current depends on the number of turns. Hence by making the

number of turns in AB very great, we can secure a very high e.m.f. for the induced current. We shall have an important application of this presently in the Induction Coil.

It will be convenient at this point to set out briefly the laws which have been established relating to electro-magnetic induction. For 'the number of lines of force through circuit' we shall use the shorter phrase 'magnetic flux through a circuit.'

Laws of Electromagnetic Induction

- (i) If for any reason the magnetic flux through a circuit is changing, an induced current will flow round the circuit;
- (ii) The induced current will continue only so long as the magnetic flux continues to change;
- (iii) The induced E.M.F. at any instant is proportional to (a) the rate at which the magnetic flux is changing at that instant, and (b) the number of turns of which the circuit consists;
- (iv) the direction of the induced current is such that its reaction tends to stop the motion which produced it.

Self-Induction. Consider the coil K in the figure, through which a current can be suddenly sent by closing the switch S. Lines of force at once make their appearance. Their directions (given by Maxwell's Corkscrew rule) are as shown in the figure. Inside the coil they are going 'into the paper.' So far as lines of force are concerned, the effect is the same as though a magnet, with its north pole facing the coil, were suddenly pushed at the coil from the front.

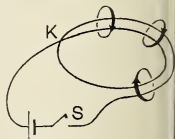


FIG. 38/6

The induction effect will be the same as though we were dealing with lines of force from this imaginary magnet—if an induced current will develop in a direction tending to drive the magnet back. Such a current would have to flow anticlockwise—opposing the current from the battery. Thus the battery current would take a longer time to reach its

maximum value, because of the opposing effect of the induced current.

By similar reasoning we conclude that if the current in K sharply declines in value, an induced current will be set up in the *same direction* as the battery current. The effect in both cases, then, is to prevent the battery current changing in strength as quickly as it would otherwise have done. The phenomenon is known as *self-induction*.

If when a current is flowing in the circuit, we open the switch very slightly, we shall probably see a small spark cross the contacts. In this case an induced E.M.F. has been set up tending to keep the battery current going, but as the circuit is broken, this 'back E.M.F.' produces the sparking discharge observed.

Induced Currents sometimes a Nuisance. Induced currents are sometimes a nuisance. Thus in the dynamo, induction coil and other instruments, the insulated wire of the coil is wrapped round a soft iron core (A.B., Fig. 38/7). When a current is started or stopped in the coil, an induced current known as an *eddy* current is set up in the core, which becomes heated. This represents a waste of energy. To prevent it, a core is made up of thin iron sheets (in the dynamo) or iron wires (in the induction coil) insulated from each other, e.g. by being varnished. There is then no continuous path presented to the eddy current.

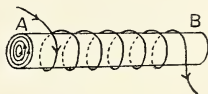


FIG. 38/7

To prevent self-induction in a coil (e.g. of a resistance box) the wire is first bent on itself, and then wound double (Fig. 38/19 on p. 484). You will be able to see for yourself why this method is effective.

The Induction Coil. You have probably seen an induction coil in operation (Fig. 38/8), even if you have not understood very much about it. It is very fascinating to see and hear the snapping sparks, like lightning on a small scale (the 'snapping' is a small edition of thunder). Sometimes, after the inventor, it is called a Ruhmkorff coil. Let us try to understand how it works.

First we have the *primary circuit*. This consists of a

battery B, connected through a switch and condenser with the *primary coil*, made up of comparatively few turns—200, perhaps—of rather thick insulated copper wire. Inside the coil is an iron core consisting of a bundle of soft iron wires varnished so as to insulate them from one another. This core greatly increases the magnetic field inside the coil when a current is flowing through the primary circuit. The insulation between the wires serves to prevent the formation of eddy currents. These would simply heat the core, wasting energy that would otherwise have been available for the current in the secondary coil.

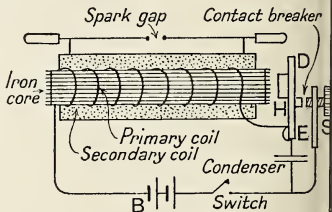


FIG. 38/8

Surrounding the primary coil, and carefully insulated from it, is the secondary coil (shown in section in the figure). This consists of many thousands of turns¹ of very thin insulated copper wire terminating in the two poles of the spark gap, of which at least one is movable.

Notice the contact-breaker. For clearness it is shown open in the figure, but it is normally closed under the action of the spring DE, the screw S being adjusted until the surfaces just touch. The contact surfaces are of platinum or tungsten. Let us start from this closed position. On closing the switch a current flows through the primary coil. The core becomes magnetised and attracts the soft iron head H. This opens the gap at the contact-breaker, the current ceases to flow through the primary coil, and so H is no longer attracted. It is carried back by the spring, closing the gap in the contact-breaker, and the whole process is repeated again and again at a rate depending partly on the mass of the soft iron head H and partly on the strength of the spring.

When the current is 'made' in the primary coil, a current

¹ Coils have been made with hundreds of miles of wire on the secondaries. Some of these give sparks of 24 in. or more, indicating a voltage of perhaps 500,000.

is induced in the secondary. The e.m.f. of this induced current depends on the number of turns in the secondary, and that is why this number is made very large. The e.m.f. in the secondary also depends on the *rate* at which the primary current is made or broken. Now the current is broken much faster than it is made, and so a much higher e.m.f. is produced at 'break' than at 'make.' It is only at 'break' that the e.m.f. is sufficiently high to enable the secondary current to jump the gap (marked 'spark gap' in the figure) provided the poles are not too far apart. Thus, though the current in the secondary is intermittent, it is not alternating.

We must now consider the condenser. It consists of alternate layers of tin foil and waxed paper (p. 420), and plays an important part. Notice that in an electrical sense it lies between the surfaces of the contact-breaker, half of the tin-foil sheets (alternate ones of course) being connected with one surface, and the other half with the other surface.

When the primary circuit is broken, an induced current in the same direction is set up (self-induction). This would spark across the gap, and quickly destroy the contact surfaces. The condenser prevents, or at least greatly reduces, this tendency, for the current takes the alternative path offered to it and charges up the condenser. So far we have two advantages:

- (1) Less damage to contact surfaces;
- (2) 'Break' more quickly effected, and therefore a higher e.m.f. induced in the secondary (and consequently a longer spark).

But while the contacts are still separated, the charge in the condenser rushes back through the primary. If we call the original primary current left-to-right, this current from the condenser is, of course, right-to-left. Now, the e.m.f. produced in the secondary by the *making* of a right-to-left current will reinforce the e.m.f. produced by the *breaking* of a left-to-right current, and this gives us advantage (3).

The induction coil was formerly used a great deal for X-ray work, and it is still employed for passing electricity through gases at low pressures. In our next paragraph we shall see

that the apparatus used for telephoning includes a simple form of induction coil. 'Coil ignition,' too, used on most motor-cars for igniting the mixture of petrol vapour and air in the cylinders is only a slightly modified form of induction coil. In this case the contact-maker is operated by a cam geared to the engine, and a distributor, also geared to the engine, controls the timing of the sparks in the cylinders. Fig

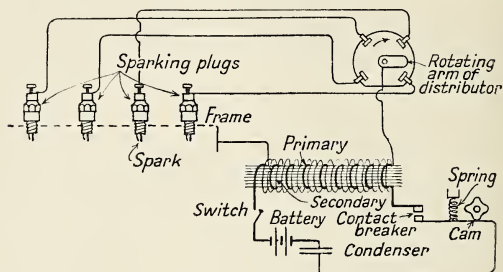


FIG. 38/9

38/9 is only diagrammatic. Actually, both primary and secondary circuits are completed through the metal frame of the car.

The Telephone. As invented by A. Graham Bell in 1876 both transmitter and receiver of the telephone took the form illustrated in Fig. 38/10. A horse-shoe magnet has soft iron pole-pieces projecting from each end, and these very nearly touch a diaphragm *D* made of thin sheet iron. Round each pole-piece is a coil of wire consisting of a great number of turns, the two coils being in series.

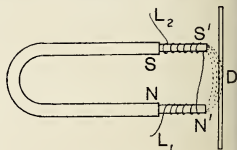


FIG. 38/10

Suppose a 'C' tuning-fork (256 vibrations per second) is sounding to the right of *D* in the figure. It sets up compressions and rarefactions of the air which cause *D* to vibrate at the same rate as the fork. Now the lines of force passing from *N'* to *S'* are highly concentrated in *D*, and each movement of *D* causes a change in the magnetic flux through the

coils. This sets up an e.m.f. which gives rise to a current in the coils, and the current passes through the line wires L_1 , L_2 to a similar arrangement (magnet, coils and diaphragm) at the receiving end. Here the variation in the current (256 times per second) will produce corresponding variations of magnetic flux in the coils, and the diaphragm will move to-and-fro 256 times per second, causing the note 'C' to be heard. We have considered only a simple form of vibration, but even in the case of very complex ones the diaphragm at the receiving end copies the movements of that at the transmitting end, and so any kind of sound (e.g. conversation) is reproduced.

The currents transmitted in the way just described are very weak, and for distances beyond 200 miles or so, at the most, the original 'Bell' telephone was found to be unsuitable, though it is still used in a slightly modified form as a receiver.

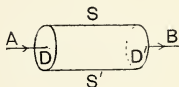


FIG. 38/11

The modern transmitter depends for its action on the principle illustrated by Fig. 38/11. DD' consists of a box filled, rather loosely, with carbon granules. D, D', the ends of the box, are carbon plates, while the sides S, S', are of non-conducting material.

When current is flowing along the path ADD'B. It is found that if D is pushed slightly inwards so that there is better contact between the carbon granules, there is an increase in the current because of the reduced resistance. Similarly,

if D moves outwards, the contact between the granules is not so good, and the current is reduced.

The transmitter is shown in section in Fig. 38/12. The carbon diaphragm D vibrates in correspondence with the speaker's voice. The granules are represented at G, and the current in the battery circuit, which includes the primary

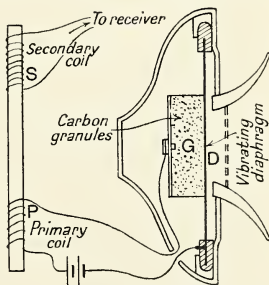


FIG. 38/12

coil P, varies with the movement of D as already explained.

P and S, wound over a soft iron core, form practically the primary and secondary coils of an induction coil, S containing many turns of thin wire. Variations of the current in P cause much greater variations of the e.m.f. in S, and these set up a varying current, which affects the distant receiver as already described. It will be seen from the figure that it will also affect the receiver of the person telephoning, and so he hears his own voice, but that produces no confusion.

A circuit extending over many miles (from transmitter to receiver) would have a very high resistance. If the carbon granules formed part of such a circuit, the variations in their resistance would be such a tiny fraction of the total that the effect would not be noticed. With the arrangement just described, however, the granules form part of a battery circuit of very small resistance, and in comparison with this the variations caused by the changing pressure of the diaphragm on the carbon granules are quite appreciable.

The Transformer. This is an instrument used for converting alternating current from a low voltage to a high one ('step-up' transformer) or from a high voltage to a low one ('step-down'). Its principle will easily be understood from Fig. 38/13, which shows a ring of soft iron, round which is wrapped a coil of many turns of insulated wire at H, and another coil of fewer turns at L. H and L form parts of two distinct circuits. An alternating current in H sets up an alternating magnetic flux in the ring, which in turn sets up an alternating current in L. It was shown in *Junior Physics* that *the ratio of the E.M.F.'s in the two coils is equal to the ratio of the respective number of turns.*

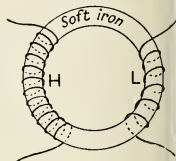


FIG. 38/13

Transformers in actual use are usually rectangular, not circular in shape, and the core is so constructed as to avoid eddy currents (p. 493). Again, the secondary coil is usually wound *over* the primary; and there are other differences.

Of the many cases in which transformers are employed we must mention one or two. For reasons which will be discussed

cussed in our next chapter, it is much cheaper to convey electric power by means of a small current at high voltage, than by a heavy current at low voltage. Hence the current produced at the generating stations, at perhaps 6000 volts, is stepped-up sometimes to as much as 132,000 volts for transmission across country by the grid system—you are sure to have seen the tall pylons associated with it.

On the other hand, the 6000 volts (say) generated at the power-station would be quite unsuitable for domestic use, so it is stepped down to 230. For operating an electric bell about 8 volts would be suitable, so we should use a transformer again to step-down the 'mains' voltage from 230 to 8.

In our next chapter we shall see that the *energy* of an electric current (expressed in *watts*) is measured by the product *volts* \times *amperes*. By the law of conservation of energy here can, of course, be no increase in this product, however we step up or step down. If a transformer were perfectly efficient, the product would remain the same, and in actual fact a good transformer will have an efficiency of about 97%. Thus vast stores of power can be transferred swiftly and silently from one circuit to a completely independent one, with a loss of as little as 3%—a truly wonderful achievement.

The Dynamo. At the beginning of this chapter we saw that by moving a magnet in the neighbourhood of a coil of wire, or by moving a coil of wire in the neighbourhood of a magnet, we could produce an electric current. Now, the movement of which we have spoken is simply a mechanical process, and so we have converted mechanical energy into electricity. What is the essential change effected by the *dynamo*.

We saw, too, that we could predict the *direction* of the electric current simply by taking for granted that in producing work must be done. We could also find the direction by applying Lenz's Law. Before going further, it will be well to discuss the direction of the current in the following special case. N and S are the north and south poles of two magnets (or of a single horse-shoe magnet.) The direction of the lines of force is indicated by dotted lines and arrows. A straight wire AB, forming part of a circuit ABC, is moved smartly

downwards across the lines of force, as shown in Fig. 38/14. What would be the direction of the current in the wire?

The answer is given by **Fleming's Right-hand Rule**, and should be carefully remembered. *Place the thumb, forefinger and second finger of the right-hand so as to be mutually at right angles to each other. Then if the thumb points in the direction of Motion, and the Forefinger in the positive direction of the lines of Force, the second finger will point in the direction of the Current.*

In Fig. 38/15 the right-hand is so placed as to give the direction of the current in the wire AB of Fig. 38/14. The direction of the second finger indicates that the current will

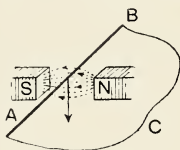


FIG. 38/14



FIG. 38/15

run from A to B. We will now consider the working of the dynamo, and we shall find this 'right-hand rule' very helpful.

ABCD in Fig. 38/16 (a) stands for a rectangular coil of wire lying between the poles N and S of two magnets. We are holding it at T, where the circuit is completed, and we are twirling it rapidly, in a clockwise direction, between the magnet poles. Everything is drawn in plan.

With the coil in the position of Fig. 38/16a, the right-hand rule tells us that the direction of the current would be along AB and CD. (Do not simply *read* this, try it for yourself. Thus in the case of CD, your right thumb must point downwards—the direction of motion, and your forefinger from left to right. You will find that your second finger is then pointing in the direction CD.)

The current will maintain the direction ABCD so long as AB is moving upwards (and CD downwards)—i.e. until the coil is in the position of Fig. 38/16b. In that position, AB and CD are not *cutting* the lines of force, but are moving along them. There is therefore no current at all.

Next consider Fig. 38/16*c*. The coil has made half a revolution from the (*a*) position. AB is now moving downwards, and CD upwards. The right-hand rule tells us that the current is now running from B to A, and from D to C (previously it was from A to B and from C to D respectively). In fact, it is in the reverse direction from that shown in Fig. 38/16*a*.

When did this change take place? Clearly, when AB began to move downwards—i.e. at the moment it was passing through the (*b*) position. So long as AB continues to move downwards—i.e. until the position of Fig. 38/16*d* is reached,

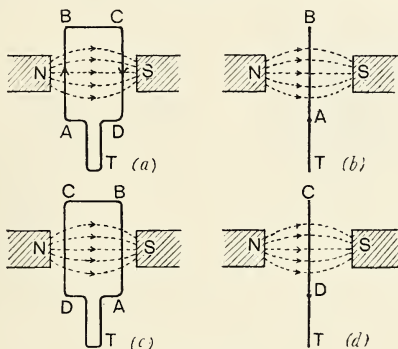


FIG. 38/16

with CD at the top—the current will retain this ‘BADC’ direction. When AB once more begins to move upwards—as soon as the (*d*) position is passed—the current will reverse once more, taking the original direction ABCD.

So far, then, the position is this. While AB is moving upwards (position (*d*), through (*a*) to (*b*)) the current has the direction ABCD. While it is moving downwards (position (*b*), through (*c*) to (*d*)), it has the reverse direction; and the change takes place each half-revolution, at the moment when the plane of the coil is at right angles to the direction of the lines of force.

Let us look more closely into the matter. Is the change

sudden or gradual? Have we a current of unvarying strength all the time that AB is moving up, then a sudden drop to zero, followed by a current of the same unvarying strength but in the opposite direction?

To answer this we recall that since the coil has a constant resistance, say R , the current is proportional to the E.M.F. ($C = E/R$), and the E.M.F. in turn is proportional to the rate at which the coil is cutting through the lines of force. Evidently, then, we have to consider this last point.

A diagram will help us. In Fig. 38/17, AD represents the coil looked at from the front (and so hiding BC). The

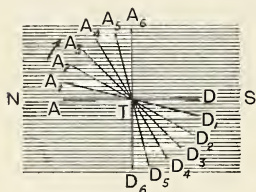


FIG. 38/17

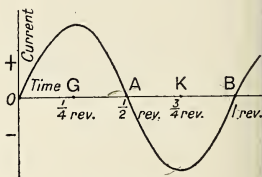


FIG. 38/18

equidistant parallel lines represent lines of force. The coil is supposed to be revolving at a uniform rate, so its position after equal intervals of time is shown by the lines A_1D_1 , A_2D_2 , etc.

By counting, we find that in moving from AD to A_1D_1 the coil has cut through ten lines of force (five between A and A_1 and five between D and D_1), from A_1D_1 to A_2D_2 through eight lines, from A_2D_2 to A_3D_3 through six, and so on through less and less, until in the neighbourhood of the A_6 position it is not cutting through any at all, but is moving parallel to them.

Thus the current is strongest when the coil is cutting the lines of force at right angles—i.e. when the plane of the coil lies in the direction of the lines of force—and it falls off until when it has moved through 90° , the current is zero, after which it reverses. The state of things is expressed graphically in Fig. 38/18, which shows how the current varies in the course of one revolution. After this the curve is repeated.

gain and again. It can be proved that the curve is a *sine curve*—i.e. that the strength of the current is proportional to the sine of the angle through which the coil has revolved from the 'zero current' position.

Such a current, changing its direction every half-revolution, is known as an *alternating current* ('A.C.'). The complete round of changes that takes place with one revolution is known as a *cycle* and the number of cycles per second is called the *frequency*. A.C. from the mains is normally '50 cycles'—i.e. it has a frequency of 50 cycles per second. Obviously the current changes in direction twice in the course of one revolution, so that the number of alternations is twice the

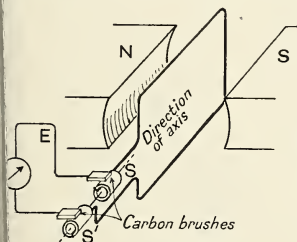


FIG. 38/19

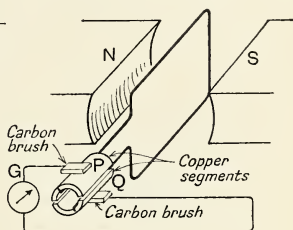


FIG. 38/20

number of cycles. The machine producing such a current is known as an *alternator*.¹

Our machine so far is not a very practical instrument. To begin with, we need some means of conveying the current to an external circuit. To do this, we have two *slip-rings*, S and S' in Fig. 38/19, mounted on the same axis as the revolving coil, but insulated from it. The two ends of the coil are connected to these slip-rings, one to each. A carbon *brush* presses against each slip-ring, and these brushes form the terminals of the external circuit E.

Suppose we want to convert our alternating current into a direct one? This is done by means of a *commutator*, a simple form of which is known as the *split-ring* commutator.

It is often called a *dynamo*, but the latter term is more correctly applied to a Direct Current Machine (also known as a *Generator*).

As in the case of the slip-rings just mentioned, it is mounted on the same axis as the revolving coil, and is insulated from it. Its two segments, P, Q (Fig. 38/20), are also insulated from one another. The two brushes previously mentioned would now press against the two halves of the commutator, one brush to each half. Each brush is so placed that just as the current is changing in direction, it makes contact with different half of the commutator.

Keep in mind that P is *permanently* in connection (by soldering) with one terminal of the coil (and, of course,

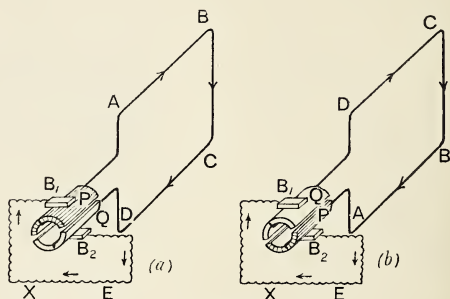


FIG. 38/21

with the other), and so every half-revolution the current that reaches P will change in direction.

In Fig. 38/21a the current is flowing from P to Q (via the path ABCD). It leaves Q by the brush B_2 , and travels through the external circuit in the direction B_2EXB_1 .

Half a revolution later we have the state of things represented in Fig. 38/21b. P is now in contact with B_2 and with B_1 . The direction of the current in the revolving coil is reversed, its direction now being from Q via DCBA to P. It leaves P by the brush B_2 , thence via EX to B_1 , and so back to Q. Thus its direction in the external circuit is B_2EXB_1 —the same as before.

But though with the device just explained we have succeeded in getting a *direct* current, we have not obtained

steady one. Representing it by a curve (OCADBE, Fig. 38/22), it is now entirely above the horizontal, its strength varying from zero (at O, A, B . . .) to a maximum (at C, D, E . . .). We could improve matters by having two coils, at right angles, so that one would be giving its maximum current while the other was giving its minimum. With this arrangement we should need four commutator segments instead of two. The currents from the two coils are shown in the figure, one with a thin, unbroken line (OCADBE) and the other with a dotted one (FGHKLM). By adding the ordinates we obtain a curve (FNCPHQ . . .) representing the *combined* current—e.g. $TX = TV + TW$. This is better than before,

but is still far from steady. The curve YZ represents the current obtained by using five coils, the individual curves, however, being omitted for clearness. There is still an obvious 'ripple' but the variation, WY, is small in comparison with the total current OW, and becomes less and less

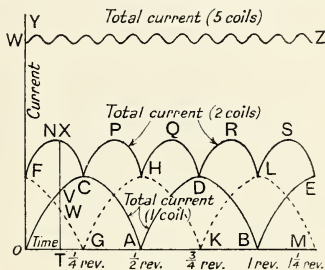


FIG. 38/22

the number of coils is increased. However, a full discussion of this part of the subject would take us too far afield.

A few words must be said about the 'winding' of a dynamo. In small instruments the 'field magnet'—i.e. the magnet producing the necessary magnetic field—consists simply of a permanent horse-shoe magnet. Usually, however, it consists of an electro-magnet, which, of course, has to be excited by passing an electric current through the *field coils*, round its poles. In some cases the winding of the field magnet is simply included in the external circuit—i.e. the entire current passes round the field magnet—and the dynamo is said to be *self-excited*. (At first, of course, there is no current, and it looks as though the dynamo would never start. Actually there is always sufficient residual magnetism in the iron pieces to set the process going.)

More commonly, alternative paths are presented to the current as it leaves the brushes. One path is via the field coils, the other constitutes the external circuit. The dynamo is then said to be *shunt wound*. The two methods of winding are illustrated in Figs. 38/23 and 24.

There is also a method of winding ('series shunt' or 'compound') which includes both the above types.

In our account of the dynamo we have constantly referred to a rotating coil consisting of a single turn. In practice, of course, the coil consists of many turns, and there are usually a number of coils. Further, in order that the lines of force cut by the coil may be as numerous as possible, we provide them with an iron path—the coils, in fact, are wound on an iron cylinder, built up, as already mentioned, of thin plates of iron, insulated from one another to prevent eddy currents. This revolving part of the instrument is known as the *armature*. However,

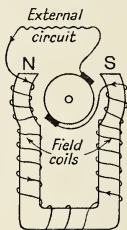


FIG. 38/23

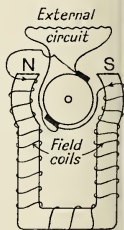


FIG. 38/24

our purpose in this chapter has been simply to explain the broad principle on which a dynamo works, and not to give a full description of the instrument as actually used.

Ammeter for A.C. It is often necessary to measure alternating current, i.e. we need a suitable *ammeter*. A hot-wire ammeter (to be described on p. 525) may be employed, or we may use what is called a *moving iron ammeter* (Fig. 38/25) in which the current passes through the circular coil shown, and in so doing magnetises two soft iron cylinders, B and F, both parallel to the axis of the coil. One of them, F, is fixed, while the other, B, is joined to a needle pivoted at P and moving over a scale. Whether the current is direct or alternating the cylinders are magnetised with like polarity, and so B is repelled by F, causing the needle to move over the scale.

Effects of Alternating Current. We should expect the

alternating current, like direct current, would produce heating, magnetic and chemical effects. The first of these is obvious enough in an electric fire, or when we handle an electric light bulb which has just been in use. For a reason which should be obvious, we can *not* show the magnetic effect by placing the 'wireless' flex over a compass needle and switching on. We can however carry out the iron filings experiment illustrated by Fig. 36/5 on p. 454. We need stout copper wire, say 2-mm. diameter, and A.C. of at least 30 amperes, previously stepped down to 12 volts or so.

To illustrate the chemical effect we can (with certain modifications) employ the apparatus represented in Fig. 34/2 on p. 424. 'S' however is replaced by a source of alternating



FIG. 38/25

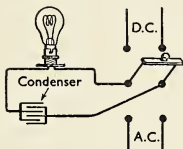


FIG. 38/26

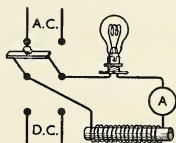


FIG. 38/27

current, either the 'mains' supply stepped down to 30 volts, or a low voltage supply stepped up to that figure. The resistance is so adjusted that the ammeter records about 0.4 amps. test-tube T_1 should be placed over E_1 (corresponding to E_2 over E_2).

We cannot now regard one electrode as anode and the other as cathode, for each is anode and cathode in rapid alternation. The result is that in each test-tube there collects a mixture of two volumes of hydrogen and one of oxygen.

Alternating Current and Condenser. Fig. 38/26 shows a lamp in series with a condenser, a switch being so arranged that either direct or alternating current can be supplied. With D.C. the lamp does not glow, but on switching over to A.C. it does. The condenser acts as a reservoir of electricity, which quickly fills up when we are using direct current, and when the current ceases. With alternating current, however,

the reservoir fills and empties with each reversal of current, and so an alternating current flows through the lamp.

The Choke Coil. With the arrangement of Fig. 38/27, either direct or alternating current can be passed through a circuit which includes a small lamp, an ammeter which will work on either D.C. or A.C., and a coil of insulated wire inside which is an iron core.

Using direct current, we find that the glowing of the lamp and the reading of the ammeter, are not affected by the presence or absence of the core. With alternating current however, the current is greatly reduced when the core is completely inside the coil, gradually increasing as it is withdrawn.

The variation is a consequence of the self-induction of the coil (p. 500). Every time the current changes direction, an opposing E.M.F. is set up, and this reduces the current just as would the insertion of a resistance. If an iron core is introduced, it becomes magnetised, giving rise to additional lines of force. When these are being produced (or, at reversal of current, withdrawn), they cut the turns of the coil, and greatly add to the opposing E.M.F. mentioned above.

The total resistance offered to an alternating current—i.e. the ordinary resistance that would be offered even to a direct current *plus* the special resistance called into play by inductance—is known as the *impedance*. In an A.C. circuit of given voltage, the current is often cut down by introducing a coil of high inductance, known as a *choke*, or *choke coil*. The resistance offered by such a coil to a direct current may be quite low.

Principle of the Motor. In the alternator or dynamo we compel a coil of wire to rotate in a magnetic field, with the result that an electric current is produced. It seems no unreasonable to expect that if we pass an electric current through a coil placed in a magnetic field, the coil will rotate. That is found to be the case, and it is this production of movement by an electric current that is the essential feature of the electric motor.

Faraday's experiment, in which, however, we have a straight wire and not a coil, is easily repeated. The arrangement of the apparatus will be clear from Fig. 38/28a. When a current is passed, the wire turns round and round (in a clockwise direction if the current is passing

down the wire as in the figure). If the direction of the current is reversed, the wire rotates in the opposite direction.

In Fig. 38/28*b* the north pole of the magnet is shown in plan, with some of the lines of force. The little circle with a cross represents the current travelling downwards. The direction of motion is given by **Fleming's Left-hand Rule**, which may be thus stated. Place the thumb, first finger and second finger of the left hand so as to be mutually at right angles to each other. Then if the **First** finger points in the direction of the lines of Force, and the **seCond** finger in the direction of the Current, the **thuMb** will point in the direction of **Motion** (Fig. 38/29).

N.B. If you find yourself confused between the Right-Hand Rule and the Left-Hand Rule, this will help. The driver of a motor (car) keeps to the left—and this *Left Hand Rule* is the *Motor Rule*.

Try this rule with reference to Fig. 38/28*b*, and you will find it leads to the conclusion that the wire should be driven round in a clockwise direction.

Now let us reconsider the case of a coil of wire between two magnet poles, but from the 'motor' point of view. In Fig. 38/16 on p. 509 we were forcibly turning the coil, and a cur-

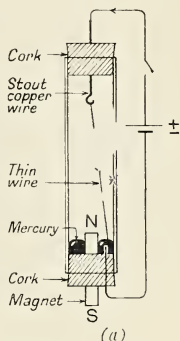


FIG. 38/28

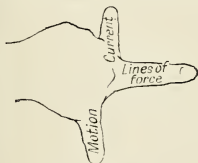


FIG. 38/29.

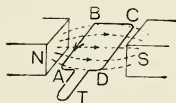


FIG. 38/30

rent was induced in it. This time we are sending a current through the coil, with a view to making it turn. If we pass the current from *A* to *B* we soon find by the Left-Hand Rule that the coil would turn anti-clockwise. Let us make it go

clockwise. The current will have to be in the opposite direction—i.e. from B to A. In what follows we shall find it more convenient to have a *front* view (Fig. 38/31*a*). B and C will be invisible, being 'covered' by A and D. The dot in A means that the current is coming through the paper *towards* the observer, while the cross in D means that it is going away from him.¹

Keeping the current in this direction, the Left-Hand Rule tells us that A would continue to move upwards until AD is vertical. By its own momentum we may suppose that it would go a little beyond that point—say, to the position shown in Fig. 38/31*b*. Now, *making no change in the direction of the current* (i.e. at A let it still be coming towards you as in Fig. 38/32*a*) apply the rule once more. You find that the

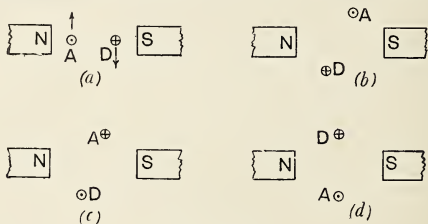


FIG. 38/31

motion would now be in the opposite direction—the coil would begin to move anti-clockwise. Clearly, if we wish it to continue moving clockwise, we must reverse the direction of the current at the moment when AD is at right angles to the lines of force—i.e. when AD is vertical; and so we get the state of affairs indicated in Fig. 38/31*c*. With this direction of current, rotation will continue to be clockwise until A has reached the bottom position (Fig. 38/31*d*). If it is to continue clockwise we must once more reverse the current.

Summing up the position, we see that if the coil is to continue moving in a given direction, the current must be reversed every half revolution, at the time when the plane of the coil

¹ Think of an arrow coming towards you. You see the *point* represented by a dot. When it is going away from you, you see the feather end, represented by a cross.

s at right angles to the lines of force. We secure this result by arranging that the current shall pass into the motor through a commutator, a simple form of which (the splitting commutator) has already been described. Speaking generally, in fact, a machine that can be run as a dynamo can also be run as a motor. Motors, too, like dynamos, are classified according to their method of winding into *series wound*, *shunt wound* and *series-shunt* (or *compound*) *wound*.

From what has been said, it would seem that to drive a motor we should need direct current (with a commutator for reversing it at each half revolution). In general this is true, but a series-wound motor may also be driven with alternating current. The reason is that in this case the same current passes through both the field coils (exciting the magnets) and the armature (the revolving coils). Thus, when the current changes direction, it changes direction in both at

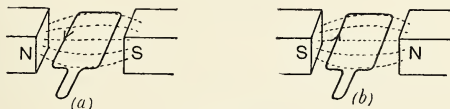


FIG. 38/32

the same time, the net result being the same as if it had not changed at all. If this is not quite clear, apply the Left-Hand Rule first to Fig. 38/32a. In Fig. 38/32b the current is reversed, causing the field magnet to have reversed polarity, and so causing the lines of force to run from right to left instead of from left to right. In both cases the Left-Hand Rule shows that the coil will be turned clockwise.

If a current is passed through a small motor and then through some lamps arranged in parallel, it will be found that as the motor gathers speed, the lamps grow dim. The reason is that when a motor is in action, it inevitably behaves like a dynamo as well as like a motor. After all, the coils of the nature are cutting through lines of force (just as in the dynamo)—it makes no difference in this respect whether the coils are driven by mechanical power or electric power. Thus Lenz's law an e.m.f. tending to stop the motion is set up. This reverse e.m.f. is known as the *back e.m.f.* of the motor,

and, of course, the faster the armature revolves, the stronger this back e.m.f. becomes. If the applied voltage is E and the back e.m.f. is e , the current at any instant is given by $C = \frac{E-e}{R}$. To begin with, e is, of course, zero. As it rises in value, C becomes less.

Thus a current which might safely be passed into a freely running motor (because it is reduced in the way just discussed) might do serious damage if passed in at the beginning, when the coils are at rest. The difficulty is overcome by having a special resistance known as a *starting resistor* in circuit with an arrangement whereby, as the motor speeds up, more and more of this resistance is cut out.

Questions

1. If you were provided with two coils of insulated wire, a galvanometer, a Leclanché cell, a magnet and some connecting wire, how would you demonstrate the elementary phenomena of electro-magnetic induction?

Describe one important application of electromagnetic induction in everyday use. *Dur.*

2. State a law for the direction of a current set up by electromagnetic induction and describe an experiment to verify the law. *Lond.*

3. State the laws of electromagnetic induction.

A coil of wire is rotated at a uniform rate about an axis at right angles to a strong uniform magnetic field. How does the E.M.F. vary with the angle between the coil and the magnetic field? State at what positions of the coil the E.M.F. is (a) a maximum, (b) a minimum.

Also show in a diagram how the direction of the induced current is related to the directions of movement and of the magnetic field.

How may the simple arrangement described above be modified to give a direct current in the external circuit? *Camb.*

4. State the laws of electromagnetic induction.

Describe a simple form of induction coil and explain, without reference to the action of the condenser, how it is possible by means of it to produce a very large voltage across its spark-gap although the source of energy is a battery of only a few cells.

Dur.

5. What is Lenz's law? Apply the law to find the direction of the induced current in a cylindrical coil of wire when a bar magnet is brought up to the coil along the axis, with the north pole towards the nearer end of the coil.

Draw a diagram of an induction coil and explain its action.

Brist.

6. Describe the construction, and explain the action, of the transmitter and the receiver of the simple telephone.

7. How could you reduce an A.C. voltage of 200 to one of 20 volts? *O.C.**

8. Describe a method of producing an alternating current.

If an alternating current supply of known E.M.F. were available, how would you transform it to one at (a) a higher E.M.F., b) a lower E.M.F.? What change would you expect to take place in the magnitude of the transformed current in each case? Give reasons. *Dur.*

9. Describe a simple alternating current generator.

Explain the difference between an alternating and a direct current and how it is possible to distinguish one from the other in practice. What is the chief advantage of alternating over direct current? *Camb.*

10. Describe, with the aid of a diagram, a simple form of series wound direct-current electric motor. Explain briefly how works.

Indicate how you would reverse the direction of rotation of the armature. *Oxf.*

11. A loose wire hangs vertically between the poles of a strong electro-magnet. State clearly, and also show on a diagram, how the wire tends to move relatively to the magnetic field when current flows in the wire. *Camb.*

12. Explain why the current which passes through the armature of an electric motor is greater when it is starting to turn than when it is turning rapidly. *Camb.**

CHAPTER 39

ELECTRICITY AS ENERGY

ONE of the prominent features of an electrical generating station in England is the large amount of *coal* that is constantly being taken in. It is used for producing steam to drive steam turbines, and these, in turn, drive the alternators ('dynamos') which generate the electric current on the principle discussed in the preceding chapter—by compelling coils of wire (enclosing cores of soft iron) to revolve against opposing forces set up by a magnetic field. In short, we have one form of energy, *heat* (from coal) converted into the *kinetic energy* of the steam turbine and of the dynamo, which in turn is converted into *electricity*. It looks very much as though electricity were simply another form of energy, and we are confirmed in this idea when we notice how it can do mechanical work (e.g. in driving the street car or operating a lift), or can give rise to heat (e.g. in the electric fire).

In Chapter 21 we found that there was a definite *quantitative* relationship between heat and work, viz. 1 British Thermal Unit = 746 ft.lb. In the present chapter we shall see that there is also such a relationship between electricity and heat, and therefore between electricity and work.

It is very easy to think of work as being measured in foot-pounds. We do a foot-pound, of course, when we lift a mass weighing 1 lb. through a vertical height of 1 ft. and if our pound weight falls a foot it does work equal to 1 ft.lb.

The work that can be done by a mill-stream (in driving a water-wheel, for instance) depends on the *weight* of water that falls and on the *difference of level* of the water before falling and afterwards. Suppose, for instance, that 11,000 lb. of water fall 5 ft. every second, the work done is $11,000 \times 5$ or 55,000 ft.lb. per second. This is the *rate* of doing work or as we call it, *the power*.

Have we anything parallel to this in electricity?

convenient unit of quantity in electricity is the *coulomb*, while *difference of potential* or *voltage* (measured in volts) corresponds to the difference of water level—the ‘fall.’ Just as a pound weight falling a foot does work equal to 1 foot-pound, so a coulomb falling in potential by a volt does a quantity of work which might perhaps have been called a volt-coulomb, but is actually known as a *joule*, a joule being equal to 10^7 ergs (p. 212).

In the ‘water’ case worked out above, the power was expressed in ft.lb. per sec. (usually converted to horse-power by dividing by 550). In the electric case, suppose we have a current of C amperes (i.e. a current which carries C coulombs per second—p. 426), and the voltage is E . Then since 1 coulomb falling 1 volt does work = 1 joule, C coulombs falling E volts does work = CE joules. That is the work done per second, i.e. it is the *power*. The unit of power (1 joule per second) is known as the *watt*. In short, then, a current of C amperes with a fall of potential of E volts has a power or rate of working of CE watts, or CE joules per second.

When we pay an electricity bill, we are really paying for the amount of *energy* consumed. This is arrived at by taking account of: (i) the *rate* at which the energy is being used (as we have seen, this can be measured in watts, but the practical unit is 1000 watts, called the kilowatt), (ii) the *time* measured in hours. These two together give us the **kilowatt hour**, defined as *the energy supplied in one hour by a circuit working at the rate of 1000 watts*.

It is very useful to be able to reckon, quickly, the cost of using electric lights, fires, immersion heaters, etc. We must, of course, know the price per ‘unit’ (i.e. per kilowatt-hour). Suppose this is a penny. Then a 100-watt lamp in 1 hour uses $\frac{1}{10}$ th of a k.-h. costing $\frac{1}{10}d.$; so it gives us 10 hours’ light for a penny. An electric fire taking 3 kilowatts would cost 3 p. per hour, and so on.

In Chapter 21, where we considered the relation between heat and work, we saw that 1 calorie = 4.2 joules, or 1 joule = $\frac{1}{4.2}$ calorie. Thus a current of C amps. at E volts, if converted into heat (e.g. by passing through a wire) would produce heat at the rate of $EC \times 0.24$ calories per second.

Example 1. A current of 5 amp. at 230 volts is used to heat a litre of water in an electric kettle. Assuming that all the heat goes into the water, and the latter is at 16°C ., how long will it be before the water boils?

$$\begin{aligned}\text{Heat required to boil 1000 gm. of water} &= 1000 \times (100 - 16) \\ &= 84,000 \text{ cal.}\end{aligned}$$

$$\begin{aligned}\text{Heat produced per sec. by electric current} &= EC \times 0.24 \text{ cal.} \\ &= 230 \times 5 \times 0.24 \\ &= 304 \text{ cal.}\end{aligned}$$

$$\begin{aligned}\therefore \text{no. of secs.} &= \frac{84,000}{304} \\ &= 5 \text{ min. } 4 \text{ sec.}\end{aligned}$$

We have seen that the power of a current in watts, or joules per second, is given by the product EC . This may be expressed in another way, for by Ohm's Law, $C = E/R$, and therefore $E = CR$. Thus instead of $E \times C$ we may write $CR \times C$ or C^2R (i.e. joules per second, or watts). Expressed in calories per second, this would be $C^2R \times 0.24$, and the heat produced in t seconds would be $C^2Rt \times 0.24$ calories.

If this expression is correct, we ought to be able to prove that the heat produced by a current is proportional to (i) the time, (ii) the resistance, (iii) the square of the current. We can do this by experiments such as the following.

A complete calorimeter is used, though only the inner vessel (covered with a slab of wood) is shown in Fig. 39/1. In the slab holes are bored and through these pass (i) a thermometer reading to fifths of a degree C., (ii) a stirrer, (iii) two pieces of stout copper wire, connected at the bottom with a coil of very thin, insulated wire, preferably made of nichrome or German silver. The coil of wire forms part of a circuit, which also includes an ammeter A , an adjustable resistance R , a battery B , consisting of two or three 2-volt accumulators connected in series, and a switch.

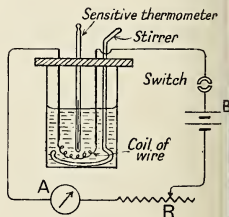


FIG. 39/1

(i) *To show that heat produced is proportional to time.*

Take the temperature of the water at intervals of 1 minute. It will be found to rise uniformly, showing that the heat produced is proportional to the time. Plotting temperature against time you should, of course, obtain a straight line.

(ii) *To show that heat produced is proportional to resistance.*

Take a coil of wire twice as long as before, cut from the same reel. This will have twice the resistance. Now repeat the last experiment using the adjustable resistance to keep the current to the same strength as before, and continuing the experiment for the same time. It should be found that the rise of temperature is twice as great. Other ratios may be tried.

(iii) *To show that heat produced is proportional to square of current.*

With a given coil inside the calorimeter carry out two experiments, adjusting the external resistance so that the current in the second experiment is twice what it was in the first. The time for each experiment must be the same, say 5 minutes. It should be found that with *twice* the current we obtain *four times* the rise of temperature, and therefore four times the amount of heat (since the mass of water was the same in each case). Ratios other than 1 : 2 may be tried, and it is a good plan to plot the rise of temperature against C^2 , when a straight line should be obtained.

Hot-wire ammeter. This depends for its action on the fact that a wire is heated when a current passes through it.

ACB is such a wire. At C it is connected with another wire CED, fixed at D. When ACB is heated by the current, it sags, and this would cause CED to become slack. But CED is connected at E with a thread which passes round the pulley P and is attached to the spring S. This thread and spring keep CED taut. The more ACB expands, the more is CED pulled on by the spring and thread, and the more is the pulley P turned. A pointer attached to P moves over graduated scale as shown in Fig. 39/2.

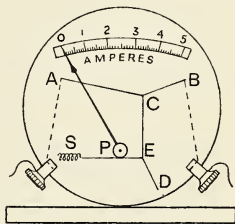


FIG. 39/2

The heating of the wire is, of course, quite independent of the direction of the current, and so a hot-wire ammeter can be used for measuring alternating current.

At this stage let us consider a few more calculations based on the formulae previously mentioned in this chapter. Many of them obviously have a practical bearing.

Example 2. Describe an experiment to show the relation between the heat generated in a current-carrying wire and the current passing through it.

A certain heating coil of resistance 50 ohms can carry a maximum current of 5 amp. Calculate (a) the number of calories of heat generated per minute in the coil when on maximum load, (b) the cost of the electrical energy used per day of 24 hours at twopence per kilowatt-hour. Lond.

For first part of question, see preceding page.

$$(a) \text{ No. of calories per second} = C^2 R \times 0.24.$$

$$\begin{aligned} \therefore \text{no. per minute} &= C^2 R \times 0.24 \times 60 \\ &= 5^2 \times 50 \times 0.24 \times 60 \\ &= \mathbf{18,000} \end{aligned}$$

$$\begin{aligned} (b) \text{ Power in watts} &= C^2 R = 5^2 \times 50 = 1250 \\ &= 1.25 \text{ kilowatts} \end{aligned}$$

$$\therefore \text{electrical energy used in 24 hours} = 1.25 \times 24 \text{ kilowatt-hours.}$$

$$\begin{aligned} \therefore \text{cost at 2d. per k.h.} &= 1.25 \times 24 \times 2 \text{ pence} \\ &= \mathbf{5s.} \end{aligned}$$

Example 3. Define the watt and the joule.

If an electric kettle contains a 1000-watt heating unit, what current does it take from 230-volt mains? How long will the kettle take to raise 1000 gm. of water at 15° to the boiling point, if 90% of the heat produced is used in raising the temperature of the water? How much would this cost if the charge is one penny for a kilowatt-hour? Camb.

For first part of question, see pp. 521 and 212.

$$\text{To find the current. } \text{Volts} \times \text{Amps.} = \text{Watts.}$$

$$\begin{aligned} \therefore \text{current (amps.)} &= \text{watts} \div \text{volts} \\ &= 1000 \div 230 \\ &= \mathbf{4.35 \text{ amps.}} \end{aligned}$$

$$\text{Heat produced per sec.} = \text{watts} \times 0.24 = 240 \text{ cal.}$$

Of this, only 90% = $\frac{9}{10}$ of 240 or 216 cal. per sec. is used in heating the water.

No. of cal. required to raise 1000 gm. from 15° C. to boiling point = $1000(100 - 15) = 85,000$.

$$\therefore \text{time required} = 85,000 \div 216 = 394 \text{ secs.} = \mathbf{6 \text{ mins. } 34 \text{ secs}}$$

Cost. 1000 watts for an hour (3600 secs.) costs 1d.

$$\therefore 1000 \text{ ,, ,, } 394 \text{ secs. costs } \frac{394}{3600} \text{d. or } \mathbf{0.109d.}$$

Example 4. Explain what is meant by the statement that the resistance of a wire is 10 ohms.

Calculate the resistance of a machine which takes 1000 watts at 200 volts. What power would be used in the machine if it were connected (a) in series with a 10-ohm resistance (b) in parallel with the resistance, and the combination were connected in each case to a 200-volt supply? J.M.B.

For first part of question see p. 468.

Resistance of machine $E \times C = 1000$

But $E = 200$

$\therefore C = 5$ amps.

By Ohm's law $C = E/R \therefore R = \frac{E}{C}$

$\therefore R = 200 \div 5 = 40$ ohms.

(a) In series with 10-ohm resistance,

total resistance $= 10 + 40 = 50$ ohms.

$\therefore C = E/R = 200/50 = 4$ amps.

\therefore power $= EC = 200 \times 4 = 800$ watts or 0.8 k.w.

(b) If the machine (resistance 40 ohms) were in parallel with a 10-ohm resistance, only $\frac{1}{5}$ th of current would pass through machine (p. 475, footnote).

Combined resistance R is given by $\frac{1}{R} = \frac{1}{40} + \frac{1}{10}$ giving $R = 8$ ohms.

$\therefore C = E/R = 200/8 = 25$ amps.

Current passing through machine $= \frac{1}{5}$ th of this $= 5$ amps.

\therefore power $= EC = 200 \times 5 = 1000$ watts
 $= 1$ kilowatt.

Example 5. A battery of E.M.F. 4.5 volts and internal resistance ohms has its terminals connected in turn by wires of resistance ohm and 4 ohms respectively. Compare the rates at which heat is generated in the two wires.

In first case total resistance (wire + battery) $= 1 + 2 = 3$ ohms

\therefore current $= E/R = \frac{4.5}{3} = 1\frac{1}{2}$ amps.

In second case total resistance $= 4 + 2 = 6$ ohms.

\therefore current $= E/R = \frac{4.5}{6} = \frac{3}{4}$ amp.

Now heat produced in wire $\propto C^2R$

$\therefore \frac{\text{heat produced in 1st case}}{\text{heat produced in 2nd case}} = \frac{\left(\frac{3}{2}\right)^2 \times 1}{\left(\frac{3}{4}\right)^2 \times 4} = 1$

i.e. heat is generated at equal rates.

Example 6. The wires leading to a group of six glow-lamps arranged in parallel have resistance 1 ohm and the lamps have each a resistance of 60 ohms. Calculate the ratio of the heat developed in the lamps to that developed in the leading wires. Dur.*

Since the lamps are in parallel, the resistance of the group will be $60 \div 6 = 10$ ohms

Leading wires have resistance of 1 ohm.

Since heat produced $\propto C^2R$ and the same current passes through the leading wires, and the group of lamps

$$\frac{\text{heat produced in lamps}}{\text{heat produced in wires}} = \frac{C^2 \times 10}{C^2 \times 1} = \frac{10}{1}$$

i.e. required ratio = **10 : 1.**

Example 7. On what factors does the heat generated in a conductor by an electric current depend? Write down an expression for it

Ten accumulators, each of E.M.F. 2 volts and internal resistance 0.5 ohm, are to be charged from 100-volt D.C. mains. Draw a diagram of the circuit used, calculate the extra resistance required to maintain the charging current at 0.8 amp., and find how much heat is generated per minute in each accumulator ($J = 4.2$ joules per calorie). Camb.

For first part of question see p. 522. Fig. 39/3 shows the circuit diagram.

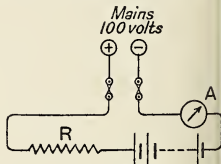


FIG. 39/3

Resistance. The accumulators provide an opposing E.M.F. of $10 \times 2 = 20$ volts.

\therefore the effective voltage for the circuit is $100 - 20 = 80$ volts

Total resistance will be given by R in the equation $C = E/R$ where $C = 0.8$ and $E = 80$.

$\therefore R = 80 \div 0.8 = 100$ ohms.

Of this, $10 \times 0.5 = 5$ ohms is provided by the accumulators.

\therefore extra resistance required = $100 - 5 = 95$ ohms.

Heat generated per minute in each accumulator is given by

$$H = \frac{C^2 R t}{4.2} \text{ cal.}, \text{ where } C = 0.8, R = 0.5 \text{ and } t = 60.$$

$$\therefore H = \frac{0.64 \times 0.5 \times 60}{4.2} = 4.57 \text{ cal.}$$

Transmission of electric power. In the preceding chapter it was mentioned that when electric power has to be transmitted over great distances (e.g. by the 'grid' system) it

first stepped-up to a high voltage. We can now see the reason for this. At the generating station, suppose the current produced has a strength of C amp. at 11,000 volts. It is stepped up, say, to 132,000 volts. There is of course no increase in power, otherwise we should have a violation of the law of Conservation of Energy; so as the voltage goes up to twelve times its previous value, the current falls from C to $\frac{1}{12} C$.

Now suppose the original current (C amp.) had been transmitted through a wire of resistance R ohms.

Heat produced per second, $h_1 = C^2 R \times 0.24$ cal.

In transmitting the stepped-up current,

Heat produced per second, $h_2 = (\frac{1}{12} C)^2 R \times 0.24$ cal.

$$\therefore \frac{h_2}{h_1} = \frac{(\frac{1}{12} C)^2 R \times 0.24}{C^2 R \times 0.24} = \frac{1}{144}$$

Thus the heat lost in transmission, using the higher voltage, only $1/144$ th part of what it would have been at the lower. Theoretically, there is another way (i.e. instead of stepping up) of reducing the loss of heat to $1/144$ th of its original value. We could do it by reducing the resistance. To do this, we should need a wire of 144 times the cross-sectional area, and the weight would be so enormous as to make the method impracticable.

Questions

A

1. An electric kettle takes a current of 5 amp. at 200 volts. What rate is it taking electrical energy?
2. Describe an electrical experiment that you have either performed or witnessed to show that 1 calorie of heat is equivalent to 4.2 joules of work.

An electric heater is marked '240 volts, 1500 watts'; calculate (a) its resistance when in use, (b) the quantity of heat produced in 5 minutes, and (c) the cost of running the heater for 5 hours, the price of electrical energy being $1\frac{1}{2}d.$ per Board of Trade unit (kilowatt-hour). *C.W.B.*

3. In the kettle of question 1, how many calories would be produced in 1 min, 40 sec.? (1 calorie = 4.2 joules).

4. A lamp is rated at 230 volts, 60 watts. What does this rating indicate? What current does the lamp take and what is its resistance? Find the cost of using this lamp for 500 hours if the cost of electrical energy is $3d.$ per unit. *Lond.**

5. A wire of resistance 10 ohms carries a current of 15 amp. How much heat would be produced in 20 sec.?

6. An alternator is maintaining a current of 10 amp. at 100 volts. By means of a transformer the voltage is stepped up to 10,000. Assuming there is no waste of energy in the process, what would be the new value of the current?

7. If in the last question the current is passing through a wire of resistance R , how much heat is dissipated per second in the wire, (i) before the process of stepping up, (ii) afterwards?

8. A lamp is marked 240 V., 100 W. What current will it take? What is its resistance? How many hours of light will it give for a penny, assuming that the cost of electrical energy is 2d. per Board of Trade unit?

9. An electric fire when full on takes 3 kilowatts. Find the cost of running it for 4 hours, at $\frac{3}{4}$ d. per kilowatt hour.

B

1. Describe an experiment which would show how the ratio of rates of production of heat in each of two coils of wire when the same current is sent through them is related to the resistance of the coils. What result would you expect?

Compare the rates of production of heat when an electric heater is used on a 220 and on a 200-volt circuit, assuming the resistance to be the same in each case. *J.M.B.*

2. An electric lamp is marked 100 W., 240 V. Calculate (a) the strength of the current passing through the filament, and (b) the resistance of the filament, when in use in the appropriate circuit. Would you expect the resistance of the filament to have the same value when cold? Give reasons.

If three such lamps are connected in parallel in a 240-volt circuit, calculate (a) their equivalent resistance, and (b) the cost of running the lamps for 75 hours when the cost of electric energy is $5\frac{1}{2}$ d. per kilowatt-hour. Give a diagram of the circuit showing the positions of lamps, switches, and fuses. *C.W.B.*

3. A current of 2 amp. passes through a resistance consisting of two coils of resistance 4 ohms and 6 ohms respectively, arranged in parallel. Calculate (a) the equivalent resistance of the coils, (b) the current in each coil, (c) the heat in calories developed in each coil in 5 minutes. How would you test the last result experimentally? *Lond.*

4. Deduce an expression for the heat developed in a conductor due to the passage of a steady electric current.

Two wires of the same cross-section have specific resistances in the ratio of 2 to 1. The wire with the larger specific resistance is three times the length of the other. Compare the heats developed in the two wires when they are placed across the main in (a) series, (b) parallel. *O.C.*

CHAPTER 40

WIRELESS AND RADIOLOCATION

THE working of the familiar wireless set depends partly on principles with which we are already familiar, partly on some that will be new to us. In the following pages we shall recall the familiar principles and endeavour to make the new ones clear; and though within our limits of space we cannot give anything like a full discussion of the subject, we may hope to give a good idea of its broad outlines.

Suppose we have some water in the U-tube of Fig. 40/1. By tilting the tube, closing the left-hand limb with a finger, and then restoring it to the upright position, we can easily

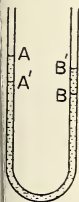


FIG. 40/1

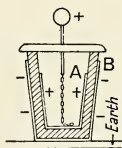


FIG. 40/2

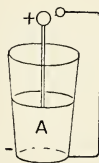
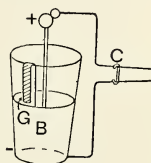


FIG. 40/3



range that the level at A is higher than at B. On removing the finger, the levels are not restored at once. The left-hand column falls, but overshoots the mark, falling to A', while B rises to B'. This 'overshooting' is repeated a number of times before the two columns are at the same level.

It has been proved that a similar situation arises when a Leyden jar is discharged (Fig. 40/2). The charged surface A is connected with the uncharged surface B. A spark is given as the discharge takes place, but the discharge is really oscillatory. There is a rush of electricity from one surface to the other, then a back-rush and so on, until the two surfaces are at the same *electric* level—i.e. at the same potential.

Next consider the experiment represented in Fig. 40/3.

A and B are two similar Leyden jars, but the inner coating of B is continued by a strip of tin-foil, and brought over the top to finish near the outer coating, the gap at G being less than a millimetre. C is a sliding contact, so that we can alter the length of the circuit between the inner and outer coatings.

A is charged, then discharged—and we watch the gap at G. Most likely nothing happens. We alter the position of C and repeat the charging and discharging. By and by we find that, *for a certain position of C*, a spark appears at G when A is discharged.

We are reminded very much of the *resonance* that may be produced by a tuning-fork. If a tuning-fork A is sounded near another one B, B will be set in vibration provided it has the same frequency as A (p. 337). In the case of the tuning-fork, the *prong* oscillates. In the case of the Leyden jar, the *electrons* are oscillating. The fork sets up sound waves in the air. The electric discharge sets up some sort of waves in space.

What is the nature of these waves?

To begin with, it was found out many years ago that they travel at the same speed as light—300 million metres per second. The difference lies in their length. Thus while light waves have a length of about 4 to 8 ten-thousandths of a millimetre (the exact length depending on the colour of the waves of which we are speaking commonly have a length of from 1 to 20,000 metres, though very much shorter ones down to a fraction of a millimetre, have been produced.

Again, when the condenser is charged, all its energy is in the electric form. Now we already know that when an electric current passes along a wire, a magnetic field is set up in the space surrounding the wire. Similarly during the first rush of the oscillatory discharge, a magnetic field is set up in the neighbourhood, and there is a moment when all the energy is in the magnetic form. When the first rush is complete the energy is once more in the electric form, and so on. Thus the oscillatory discharge corresponds to the alternate production of electric and magnetic fields, and the waves given off are known as *electromagnetic waves*. Fig. 40/3a will help to indicate the similarity of these waves to those which produce light of different colours, and to the infra-red and

ultra-violet waves.¹ It shows also that X-rays and gamma-rays belong to the same family (so do the rather mysterious cosmic rays).

It would be quite impossible by using an ordinary scale to show all these waves on the same diagram. Suppose for instance that to represent the light waves—the visible part of the spectrum—we used a scale in which a micron is represented by 1 cm. (a micron = 1/1000th mm., and light waves vary in length from about 0.4 to 0.8 microns). The entire range of gamma rays would be invisible because they would be represented by only 1/700th mm., while to show the waves commonly used in wireless we should need a length of 6,000

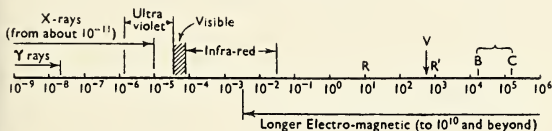


FIG. 40/3a

miles or so. However, by using a *logarithmic* scale, i.e. one in which our scale is proportional to the logarithm of the wave-length—we are able to make a very fair representation.

A wave-length of about 6 metres (V in Fig. 40/3a) is used for television. For broadcasting, the B.B.C. uses wave-lengths ranging from 206 to 1,500 metres (BC in figure), while wave-lengths of from about 10 to 1000 cm. (RR') are employed in radar.

At various parts of the scale we notice overlapping. Thus the shortest electromagnetic waves are exactly similar to the longest infra-red, though the former are produced by means of a triode valve, while the latter are present in radiation from the sun.

To return to our experiment with the Leyden jar. This shows that the waves thereby produced can set up corresponding waves in a neighbouring circuit provided the latter has been

¹ For infra-red waves see pp. 199, 200. Ultra-violet rays are very active chemically. Thus they rapidly affect a photographic plate, and their action on the skin cause 'vitamin D' to be produced just below

Further, they cause *fluorescence*—we have an example in the curious bluish appearance of paraffin oil. The oil absorbs the very short waves and gives out longer ones—long enough to give a colour effect.

suitably *tuned*—and here we have the underlying principle of ‘wireless.’

In the experiment of Fig. 40/3 most of the energy of the discharge is used up to furnish the heat, light and sound of the spark, and very little is available for furnishing electromagnetic waves. The path of discharge should include an appreciable *inductance*—a coil in which a magnetic field is set up as the current passes through it. Again, Marconi showed that much better effects are obtained if one plate of the condenser is connected to an ‘aerial’ and the other to the earth. For our oscillatory discharge much better results would be obtained by using, instead of a Leyden jar, a high-frequency alternator in conjunction with a transformer. These ideas are embodied in Fig. 40/4. The two parallel lines represent a condenser. This allows a good supply of electric energy to be stored up in our transmitting circuit. The arrow represents provision for tuning.

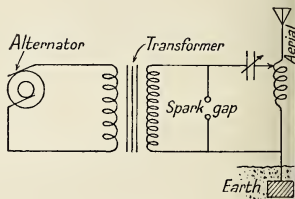


FIG. 40/4

Our ‘receiving’ circuit would be similar, except that, of course, we should have no need for the alternator and transformer. It can be shown that the frequency with which such a circuit ‘resounds,’ so to speak, is equal to $\frac{1}{2\pi\sqrt{LC}}$, where L and C represent the

inductance and capacitance respectively, expressed in suitable units. We mention this formula only as a reminder that we can alter the frequency by altering either the inductance or capacitance, or both. Usually we alter the latter, the condenser concerned being usually of the form shown in Fig. 33/23 (p. 420).

So far we have mentioned only one means of finding whether waves have been received by the second circuit—the production of a spark (as at G in Fig. 40/3). We need something much more sensitive. Could we not make the current, as it oscillates in the second circuit, operate a telephone?

Unfortunately, no—not with the arrangement as so far described. The oscillations are far too rapid to affect the comparatively slow-moving diaphragm of the telephone.

Let us examine these oscillations a little more closely. As we have seen, they accompany a spark, and gradually die away. In Fig. 40/5, A represents the oscillations accom-

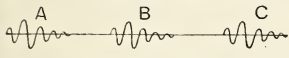


FIG. 40/5

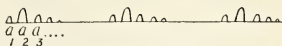


FIG. 40/6

panying a single spark. The fact that the electrons move first in one direction and then in the other is indicated by drawing part of the wave above the line, the following part being below it. The alternations are shown as gradually dying away to zero—we say they are ‘damped.’ The complete figure represents the effect of three successive sparks.

Now it is possible (by means which we will consider presently) to allow the current to pass in one direction only, so that instead of Fig. 40/5 we have Fig. 40/6. Accompanying

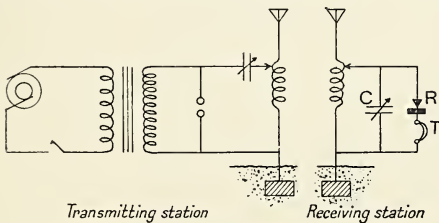


FIG. 40/7

a single spark we now have a series of pushes at the diaphragm ($a_1, a_2, a_3 \dots$), varying in intensity, but all in the same direction. We may therefore regard them as merging into a single push. The frequency of these single ‘pushes’ is therefore equal to the frequency of the sparks, and this is *not* too high to produce vibration of the diaphragm. Thus if the sparks were produced at the rate of 256 per second, we should have a note of ‘middle C’ pitch in the telephone.

If in the circuit which includes the alternator, etc.—the transmitting circuit—we include a key, we can start and stop the telephone note in the receiving circuit at will, and we could obviously use a Morse code to transmit messages. The complete arrangement is represented diagrammatically in Fig. 40/7, C being a condenser of adjustable capacity, T a telephone, and R an arrangement for 'rectifying' the current, i.e. for allowing it to pass in one direction only.

We must now take up this question of how the current is rectified. It was formerly done by inserting a suitable crystal in the circuit. Certain crystals—galena, for instance—when touched by a fine metal point, show a conductivity much greater in one direction through the point of contact than in the reverse direction, so that for practical purposes they allow the current to pass in one direction only. Such an arrangement is represented diagrammatically by R in Fig. 40/7. Instead of a crystal, however, a valve is now almost universally employed. We will first explain the principle of the *diode valve*, invented by the late Sir Ambrose Fleming.

Its action depends on the fact that when a metal filament is heated, it throws off electrons. In practice a tungsten filament F is used, surrounded by a copper cylinder P (Fig. 40/8a), both filament and cylinder being contained in an evacuated bulb. For diagrammatic purposes the cylinder (or 'plate') is usually represented as in Fig. 40/8b.

Now suppose P forms part of an *oscillating* circuit. It will be charged positively and negatively in rapid alternation.

- (i) When it is positively charged, the electrons (negative electricity) thrown off from F will be attracted. But a flow of negative electricity from F to P corresponds to a flow of positive electricity from P to F (Fig. 40/9). Thus we have a current in the direction PF—the gap serving as a conductor because of the electrons present in it.

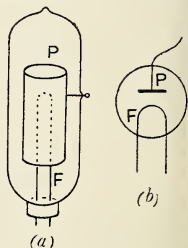


FIG. 40/8

- (ii) When P is negatively charged, the electrons from F will be repelled, and so there will be no current.

Thus what would have been an oscillating current is converted by the valve into a unidirectional current—from P to F.

As we have seen, the arrangements so far described would serve for the transmission of signals such as Morse, but would certainly not transmit conversation or music. For one thing, the sounds of conversation or music are continuous, and it is difficult to see how they could be reproduced by discontinuous waves such as we have described.

However, methods have been devised for producing continuous electromagnetic waves such as those represented in Fig. 40/10a. They are, of course, of radio frequency, and are known as *carrier waves*.

The line HL joining the crests of the waves is horizontal. By means which we cannot discuss here, it is possible to combine such waves with sound-waves, so that the line joining the crests is the wave form of the sound we are producing. Thus a tuning-fork produces a wave of very simple form (Fig. 40/10b on p. 535), and when this is combined with the waves first mentioned, we get a result such as is represented in Fig. 40/10c, except that the carrier waves should be much closer together. When a wave of this sort is rectified by cutting out the lower half, it affects the loud speaker in such a way as to reproduce the original sound. The sound waves of conversation or music are often very complex indeed, but they are imposed on the carrier waves in the same way.

Even in such a brief discussion as this, we must give some account of how waves which would be too feeble to affect a wireless set are *amplified*. For this a triode valve, shown diagrammatically in Fig. 40/11, may be used.

The special feature of this valve is the *grid*, a third electrode coming between the plate and filament of the valve previously described. In practice it consists of fine copper gauze, or a spiral of thin wire.

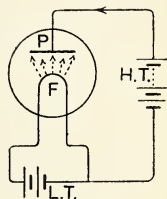


FIG. 40/9

It acts in this way. Suppose we have a plate current passing in the direction PF, as already described. The grid is made part of the circuit in which the feeble oscillating current is flowing. The potential of the grid is therefore varying all the time, though only slightly.

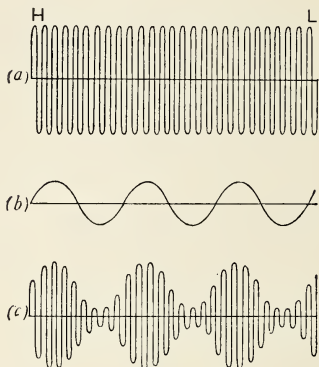


FIG. 40/10

Now, when the grid has a slight positive potential, the flow of electrons from F to P is greatly encouraged, and so there is a great increase in the strength of the plate current. When it has a slight negative potential, the flow of electrons is

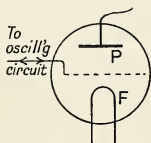


FIG. 40/11

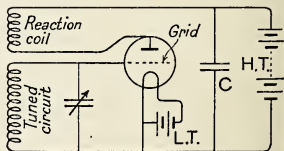


FIG. 40/12

greatly reduced, and so there is a marked decrease in the plate current. Thus, instead of the feeble variations of the original current, we have a plate current showing strong variations. In short, our triode valve has acted as an *amplifier*.

The Valve Transmitter. Consider the circuit of Fig. 40/12. Here the plate current is made to pass through a *reaction coil*, and a weak alternating current is supposed to be passing through the tuned circuit. This weak current causes slight changes in the potential of the grid, which in turn gives rise to much greater changes in the energy of the plate current. As the latter passes through the reaction coil, the current in the tuned circuit is increased (by induction). This is now able to produce greater changes than before in the strength of the plate current, and so on—the action is cumulative.

By reducing the distance between the reaction coil and the tuned circuit ('increasing the coupling' is the usual phrase), we can still further increase the effect just described. Now, an increase of current in the tuned circuit is the same result as would follow from a reduction of its resistance (Ohm's Law). By suitably adjusting the coupling, we can reduce this effective resistance to zero, and a current set up in the tuned circuit will now circulate indefinitely, since there is no resistance, so to speak, to stop it. Such a condition is known as *oscillation* (the chief cause of 'howling' in a wireless set).

By adjusting the condenser (capacity), or the inductance, we can alter the frequency of the tuned circuit. Thus we have the means of generating an alternating current of adjustable frequency, which is just what we need for transmitting. The transmitting aerial has only to be connected to the tuned circuit, and the alternating current from the latter then provides the energy which is transmitted into space. Valve transmitters based on the principles just briefly described are used by all broadcasting stations.

Having discussed the use of the triode valve as an amplifier for transmission purposes, let us now go back to the point (p. 538) at which we saw that it could be used as an amplifier in the receiving set. This amplified current may be still further amplified by another valve, and so on, until finally it enters the moving coil of a

Loud Speaker (Fig. 40/13). This consists of a very light coil wound on a short cardboard tube attached to a paper cone, and it moves between the poles N, S of an electromagnet. The direction of the current in the coil is evidently at right angles to the lines of force between N and S, so the coil moves

in and out in accordance with Fleming's left-hand rule. As a result the paper diaphragm, or cone, gives vibrations corresponding to those which originally struck the microphone at the transmitting end, and so the original sounds are reproduced.

Radiolocation. We have seen that 'wireless' waves differ from light waves only in the matter of length, and, as we should expect, they are subject to the same laws of reflection and refraction. An important method for finding the position of distant objects, aeroplanes, for instance, is based on the reflection of electromagnetic waves, and is known as *radiolocation* or *radar*.

The transmitter and receiver are in this case in the same place, and usually have the same aerial, or 'antenna,' as it is called in radar. The antenna can be moved easily into any desired direction. Suppose it is pointed at an aeroplane, and the 'modulator' is turned on (this causes the radio-frequency oscillator to oscillate violently for a millionth of a second or so).

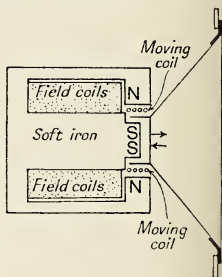


FIG. 40/13

Waves are thus sent out, strike the aeroplane, and are reflected to the receiver. This contains a device by which the time required for the to-and-fro journey is automatically recorded with an accuracy of about one-thirtieth of a millionth of a second. Suppose the time in this case is $1/100,000$ th part of a second. Now, as the waves travel 186,000 miles per second, they would travel 1.86 miles in the $1/100,000$ th part of a second. This is, of course, the distance to the aeroplane and back again, so the required distance is 0.93 mile. (The observer does not work this out—a spot of light moves to the correct number on a scale.)

Normally, of course, the aeroplane would not be visible at all. The antenna is moved about until a signal (known as a 'pip') is observed on the receiver. The direction of the object is now given by that of the antenna, and the distance is recorded as already described. The time is so short that

the aeroplane will not have made any appreciable movement, and gunfire can be as effective as if it were visible.

The transmitter does not send out waves continuously, but in very short bursts of a millionth of a second or so, followed by comparatively long intervals of rest (a few thousandths of a second), during which the receiver is in action. If very long waves were used, the instrument would not be able to distinguish between two objects very close together, so the aim has been to use wave-lengths as short as is practicable.

It is quite possible that radar turned the scale in the battle of Britain (1940), when our margin of safety was certainly a very narrow one. Without it, our pilots would have had to be in the air all the time to be ready for a possible attack, and they would have been worn out. As it was, they could remain grounded and resting, knowing that they would receive timely warning of an attack.

Radar was much used by the Army and Navy (especially the latter) during the war, as well as by the Air Force. For instance, 'On the evening of 4 November 1942 . . . the sea battle for Guadalcanal was in its final phase, the issue still undecided. Aboard the American vessel, radar, like an invisible searchlight, probed the darkness and discovered the presence of an enemy vessel more than eight miles away. The big ship lifted its gun muzzles towards the stars. They flashed and thundered. The second salvo, despite both darkness and extreme range, landed squarely on the target, which disappeared from the radar screen.' ¹

Happily, the uses of radar are not limited to the operations of war. A captain, for instance, near a fog-bound coast can 'see' the contour of the coast. He can ascertain the position of other ships in the neighbourhood, and of rocks, buoys, lighthouses, etc. He can steam ahead almost as confidently as though the sun were shining in a clear sky.

¹ From 'Radar; a Report on Science at War.'

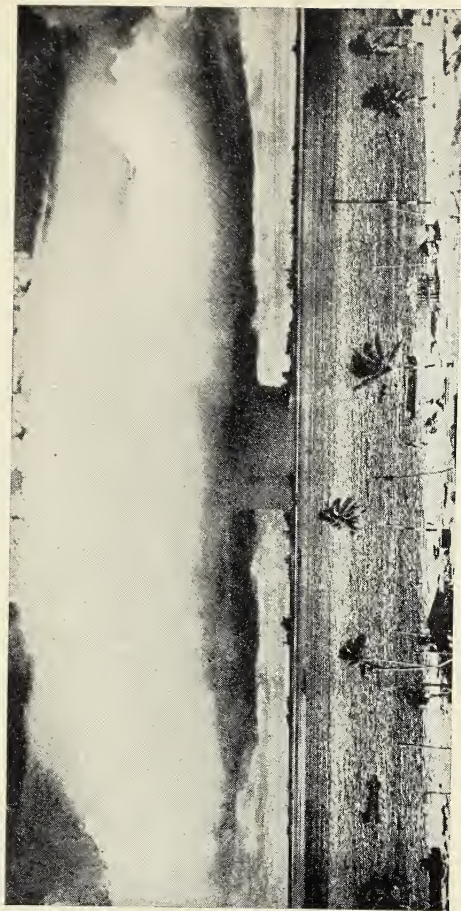
CONCLUSION

ATOMIC ENERGY

MOST people study Physics for a long time before they realise that what they are studying is chiefly *energy* in its various forms. Clear examples of this are seen in Electricity, Heat and a large part of Mechanics. In England our main source of energy is coal, and to a much less extent oil. The same is true for many other countries, though some (e.g. Norway) make great use of water-power. The energy required for blasting in quarries and coal-mines is obtained from explosive compounds, which again have their origin, partly or completely, in coal; and the same is true of the explosives used in war.

But on Aug. 6th, 1945, the world was startled by reading of the unparalleled havoc caused by a single bomb which fell at Hiroshima in Japan, causing the almost complete destruction of a large town and killing 80,000 people—more than had been killed in all the towns of Britain by all the air-raids of the war. Its destructive energy was derived, it was explained, not from any of the usual sources, but from the disruption of atoms of uranium; it was the Atom Bomb. Later, for experimental purposes, an atom bomb was dropped at Bikini in the South Pacific, to explode under water, and the illustration gives some idea of the remarkable visible effect produced. In this chapter we shall try to understand something of the nature of this new source of energy.

On p. 402 it was explained that atoms are the ultimate particles of which elements are composed, each element having its own special kind of atom. We saw, too, that these atoms are incredibly small. Until 1898 people thought of them as being something like tiny billiard balls—i.e. something 'the same all through'—without any particular structure. In that year, however, there was discovered an element, radium, which behaved in a very peculiar way. It was constantly throwing off 'alpha-particles,' really atoms charged with positive electricity, of another element called



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EXPLOSION OF ATOM BOMB AT BIKINI

helium. In addition, it was throwing off negatively charged particles called electrons.

Our first thought is, of course, 'Then radium cannot be an *element*; it must be a compound of helium.' But though it would take us too long to consider the evidence, it is quite certain that radium *is* an element, and, like any other element, is made up of atoms. The fact is that the atoms of radium are constantly undergoing a kind of decomposition that is known by the special term of *disintegration*. The atom of radium, then, is *not* 'the same all through.' Almost infinitely tiny as it is, it is built up of still tinier parts, which include electrons and positively charged atoms of helium.

Another very important point arises in connection with the disintegration of the radium atom. Much *energy* is released. It is found that the alpha-particles of which we have spoken are given off with a velocity of something like 20,000 miles per second, and as these tiny but enormously swift projectiles strike the surrounding medium, they are slowed down or stopped, their kinetic energy being converted into heat. One gram of radium will produce about 140 calories per hour. The radium is gradually used up in the process—converted into helium and other products—but the process is so slow that it would take about 1700 years for *half* of it to disappear. It is not a very difficult problem in arithmetic to calculate the amount of heat produced by the disintegration of one gram of radium. It works out to about 2640 million calories, enough to raise the temperature of 31 tons of water from room temperature to boiling point. Of course this heat is given out over a very long period of time but when we recall that it is produced from a quantity of radium weighing very much less than a sixpence, the amount is astonishing.

However, let us leave details on one side and get a clear grip of the essential point, which is that *by the disintegration of atoms of radium, enormous quantities of energy are released.*

The fact that the radium atom was constantly undergoing the change just discussed made it certain that it must possess a complicated structure—it was no mere 'billiard ball.' Men soon began to suspect that *all* atoms had a structure following the same general design, and after much patient work they

had by the year 1932 arrived at the ideas which have already been outlined on p. 402, but which we must now consider in rather more detail.

In our earlier chapter something was said about protons and neutrons. The mass of each of these is about equal to that of a hydrogen atom. The proton carries a unit of positive charge, but the neutron is uncharged. It may be that it is really made up of a proton firmly combined with an electron, the positive charge of the one cancelling out the negative charge of the other.

The nucleus is made up of at least these two kinds of particles, protons and neutrons. The nucleus of the helium particle, for instance, consists of two protons (P in Fig. C/1) and two neutrons (N), with a total mass of four (calling the mass of the hydrogen atom one). This nucleus, because of its two protons, carries two positive charges. The helium atom as a whole, however, is electrically neutral, and this neutrality is secured by two electrons (e), whose two negative charges balance the two positive ones. Fig. C/1 represents the state of affairs diagrammatically, except that the distance from the nucleus to the electrons should be very much greater.

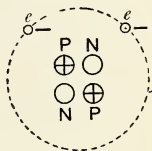


FIG. C/1

The number of electrons surrounding the nucleus varies from one atom to another. The hydrogen atom includes only one electron, the helium atom two (as already mentioned), the lithium atom three and so on, until the largest number is found in the case of uranium which has ninety-two. The number of electrons is known as the *atomic number* of the atom in question. Thus the atomic number of helium is two, while that of uranium is 92.

The *atomic weight* of an element is numerically equal to the sum total of the neutrons and protons in the nucleus, the mass of one of these being taken as the unit. Thus helium has an atomic weight of four. The mass of an electron is only about 1/1800th part of the mass of a proton (or neutron), and so in comparison may be neglected. From what has been said we should expect all atomic weights to be whole numbers, since the nucleus must contain a whole number

of protons and neutrons. If we look at a table of atomic weights, we shall find that though many of them are whole numbers, this is by no means true of all. Chlorine, for instance, has an atomic weight of 35.47. The reason is that there are really two different atoms included under the name chlorine. In one there are 17 protons and 18 neutrons giving an atomic weight of 35, while in the other there are 17 protons and 20 neutrons, giving an atomic weight of 37. Ordinary chlorine is a mixture of these two *isotopes* (as they are called), in such proportions as to give an atomic weight of 35.47. The two 'chlorines' are chemically identical. That is because each has the same number of electrons, for it is the number of electrons in the atom of an element which determines its chemical properties. We shall see presently that ordinary uranium consists of a mixture of *three* isotopes with atomic weights of 234, 235 and 238 respectively. In each case there is the same number of electrons (92), so all three have the same chemical properties. Further, the nucleus in each case contains 92 protons, the positive charge on which balance the negative charges on the electrons. The remainder of the atomic weight is, of course, contributed by the neutrons. Thus 'uranium-234' contains $234 - 92$ or 142 neutrons, while uranium-235 contains 143, and uranium-238 contains 146.

The neutron was not recognised until 1932, though it had been produced several years earlier (without being recognised) by mixing radium bromide with a rather rare metal known as beryllium. Apparently alpha-particles thrown off by the radium struck some of the nuclei of the beryllium atoms knocking out one or more neutrons from each. The neutron is, of course, very small, but it may have a velocity of some thousands of miles per second; so scientists thought it would be very interesting to use these neutrons as projectiles with which to bombard the nuclei of other atoms. They had methods (which we cannot discuss here) of recognising when they had 'made a hit.' They found that sometimes the projectile buried itself in the nucleus which it had struck but in certain cases it broke the nucleus into two or more fragments, other neutrons being set free in the process. But there was something far more important than the mere

breaking of the nucleus, and that was the release of energy (chiefly in the form of heat) which accompanied it. In the case of uranium this release of energy was to have very remarkable consequences.

Let us put it in this way. Suppose a quantity of uranium is divided into sections A, B, C, . . . Z (Fig. C/2). A neutron, say from the radium-beryllium mixture, strikes A and dislodges two fresh neutrons. We now have a total of three. Before reaching B one of them may be lost, perhaps escaping at the sides, or perhaps being absorbed in a nucleus like a bullet sticking in a clay bank. However, while only one neutron struck A, we have two striking B, and these dislodge (say) four fresh neutrons. Four plus two = six. Allowing for wastage, suppose four neutrons strike C, causing eight to strike D, and so on. Before long, the number of neutrons striking a section is enormous, and *the release of energy is correspondingly great.*

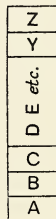


FIG. C/2

What we have sketched out is known as a *chain reaction*. Obviously it is very similar to a spreading fire, and it may get completely out of hand.

Before going further, let us grasp the point that a chain reaction from the arrangement indicated in Fig. C/2 is by no means inevitable. It might be that of 100 neutrons released from A, after allowing for losses by escape, absorption, etc., only 90 would strike B (a loss of 10%), and if this rate of loss continued only 81 would strike C, and so on. Instead of the fire spreading so to speak, it would gradually die out. Now let us see what happened in practice.

Uranium 'fission,' as it is called, was well known to prominent physicists all over the world as early as the beginning of 1939, at a time when war-clouds were already gathering. What was *not* known was whether a chain reaction could be established. If it could, there was the possibility of constructing a bomb whose destructive energy would far exceed that of any existing high explosive. Thus while British and American scientists were at work on the problem, they were well aware that German scientists were engaged on it, too. Defeat in the laboratory might well bring national downfall in its train.

Research on an enormous scale was instituted, and we may notice one or two of the problems to be solved.

It has already been mentioned that uranium as extracted from natural ores consists of a mixture of U-234, U-235 and U-238. Of these, it is only the U-235¹ that is readily fissionable, and commercial uranium contains only about one part in 140 of this particular isotope. When we recall that there is no difference in *chemical* properties between one isotope and another, we begin to see how difficult it must be to extract U-235 from the mixture.

Before attempting to make a bomb, it was necessary to find whether a chain reaction was possible. At this stage, precautions had to be taken against the possibility that the reaction, if it occurred, might get out of hand. One precaution was to arrange for a slowing down of the neutrons. This was done by the use of a very pure carbon as a 'moderator,' the 'slugs' of metallic uranium being embedded in the carbon. The neutrons struck the nuclei of the carbon atoms and were slowed down. In addition, arrangements were made for *absorbing* some of them if necessary. For this purpose, bars of cadmium were used, which could be drawn out or pushed in according as a more or less vigorous reaction was desired. The *pile* (as the whole arrangement was called) did not at first give a chain reaction, but by increasing the proportion of U-235, and adjusting the arrangements for moderating and absorbing, success in this respect was achieved.

To be effective, a bomb must not be below a certain minimum size (known as the 'critical size'). The reason is that the smaller the size, the greater is the area in comparison with the volume. Now, a large area favours escape of neutrons, and in a bomb of less than critical size the proportion of escaping neutrons is so great that a chain reaction is not possible. No information has been published about critical size, except that it apparently lies between 1 kilo and 100 kilos.

Let us now turn from our discussion of the atom bomb and consider the far more important question of whether atomic energy can be used as a source of power for ordinary peace-time purposes.

¹ U-238 however (which forms 99.3 per cent. of purified natural uranium) may be converted into plutonium, which is fissionable.

One obvious difficulty has just received mention—the unit must not be of less than a certain size. Further, a chain-reaction pile gives off rays which would quickly prove fatal to anyone receiving them, and so far it has been necessary to surround the pile with a shield of concrete 5 ft. thick. Still another difficulty is that the energy from the pile would probably in the first instance be used to heat water, which would soon become dangerously contaminated. These are just a few of the difficulties that must be surmounted before atomic energy can be utilised as an everyday source of power, and they suggest that the first units are likely to be large ones, serving perhaps as central generating stations or providing the source of power for a large works. Possibly too, an ocean liner might be driven by atomic energy. It seems unlikely however that it will ever be possible to use such energy in cases where it would have to be produced on a very small scale, say for driving a motor car or even a locomotive.

But even taking into account all the practical difficulties, it will surely be no very long time before man will have made atomic energy his main source of power. The world's great industrial nations will then not necessarily be those possessing great natural supplies of coal, oil or water-power, but those best served by their scientists and technologists—and in this respect our own outlook is a very hopeful one. Man will look back upon the days when he hewed his coal from grimy pits much as we look back upon the Stone Age, for he will have at his disposal sources of power, clean in operation and unlimited in extent, which may well provide the physical basis for a development and culture as yet undreamed of.

But at the same time he would be living in a fool's paradise if he did not contemplate the ghastly alternative, and take steps to obviate it. The power that, rightly used, can enable the human race to advance towards new heights of achievement and culture, can also reduce it to the lowest depths of misery, and may come within measureable distance of exterminating it altogether. The discovery of how to release atomic energy was a great achievement, but if it is to be a blessing and not a curse it must be followed by a still greater one—the devising of means whereby all the nations of the world may live at peace one with another.

ANSWERS TO NUMERICAL EXERCISES

Answers to the odd-numbered questions only have been given

Chapter 32, p. 397

A

1. (i) 1 dyne; (ii) 1 dyne; (iii) 4 dynes. 3. 3 dynes.

B

1. 4 cm. 3. (a) 225 webers; (b) 450 units (of moment) per c.c.
5. 1 : 2. 7. 0.235 oersted. 9. 0.181 oersted.
11. 0.195 oersted. 13. 0.193 oersted.

Chapter 33, p. 421

9. 4 dynes.

Chapter 34, p. 439

1. 0.148 gm. 3. 1.22 amps. 7. 1.35 amps.
9. 4 hr. 11. 0.0000104; 0.000328. 13. 0.0011 amp.
15. Copper gains 0.178 gm., zinc loses 0.183 gm.

Chapter 35, p. 450

3. 0.614 gm.

Chapter 36, p. 465

B

5. 0.57. 7. 0.286. 9. 36° .

Chapter 37, p. 489

1. 1.43 amp.; 1.07 amp. 3. 32.07 m. 5. 0.75 ohms.
7. (a) 2.4 ohms; (b) 1.2 amps. in 4-ohm coil, 0.8 amps. in 6-ohm. 9. 3 ohms.
11. (a) 2 volts; (b) 12 volts; (c) 1.2 amps. in 1st (10-ohm), 0.8 amps. in 2nd.
13. (a) $3\frac{2}{3}$ ohms; (b) 0.41 amp.; (c) 0.27 amp. (1-ohm branch), 0.14 amp. (2-ohm branch).
15. $\frac{1}{2}$ ohm. 17. (a) 0.68 amp.; (b) 0.68 volt; (c) 0.41 amp.
19. 3.2 ohms. 21. Put resistance of 98 ohms in series with it.
23. (a) 4.2 volts, 15 ohms; (b) 1.4 volts, $1\frac{2}{3}$ ohms.
25. 0.3 amp.; 1.2 volts.
27. (a) 0.253 ohm and 1.500 ohms; (b) 0.289 volt.
29. 1.5 volts, 0.5 ohm.

Chapter 39, p. 527**A**

1. 1 kilowatt. 3. 23,810. 5. 10,800 cal.
7. (i) 24R cal.; (ii) 0.24R cal. 9. 9d.

B

1. 1.21 : 1. 3. (a) 2.4 ohms; (b) 1.2 amps., 0.8 amp.;
(c) 415 cal., 276 cal.

	0	1	2	3	4	5	6	7	8	9	1	2	3	4	5	6	7	8	9
10	•0000	0043	0086	0128	0170	0212	0253	0294	0334	0374	4	8	12	17	21	25	29	33	37
11	•0414	0453	0492	0531	0569	0607	0645	0682	0719	0755	4	8	11	15	19	23	26	30	34
12	•0792	0828	0864	0899	0934	0969	1004	1038	1072	1106	3	7	10	14	17	21	24	28	31
13	•1139	1173	1206	1239	1271	1303	1335	1367	1399	1430	3	6	10	13	16	19	23	26	29
14	•1461	1492	1523	1553	1584	1614	1644	1673	1703	1732	3	6	9	12	15	18	21	24	27
15	•1761	1790	1818	1847	1875	1903	1931	1959	1987	2014	3	6	8	11	14	17	20	22	25
16	•2041	2069	2095	2122	2148	2175	2201	2227	2253	2279	3	5	8	11	13	16	18	21	24
17	•2304	2330	2355	2380	2405	2430	2455	2480	2504	2529	2	5	7	10	12	15	17	20	22
18	•2553	2577	2601	2625	2648	2672	2695	2718	2742	2765	2	5	7	9	12	14	16	19	21
19	•2788	2810	2833	2856	2878	2900	2923	2945	2967	2989	2	4	7	9	11	13	16	18	20
20	•3010	3032	3054	3075	3096	3118	3139	3160	3181	3201	2	4	6	8	11	13	15	17	19
21	•3222	3243	3263	3284	3304	3324	3345	3365	3385	3404	2	4	6	8	10	12	14	16	18
22	•3424	3444	3464	3483	3502	3522	3541	3560	3579	3598	2	4	6	8	10	12	14	15	17
23	•3617	3636	3655	3674	3692	3711	3729	3747	3766	3784	2	4	6	7	9	11	13	15	17
24	•3802	3820	3838	3856	3874	3892	3909	3927	3945	3962	2	4	5	7	9	11	12	14	16
25	•3979	3997	4014	4031	4048	4065	4082	4099	4116	4133	2	3	5	7	9	10	12	14	15
26	•4150	4166	4183	4200	4216	4232	4249	4265	4281	4298	2	3	5	7	8	10	11	13	15
27	•4314	4330	4346	4362	4378	4393	4409	4425	4440	4456	2	3	5	6	8	9	11	13	14
28	•4472	4487	4502	4518	4533	4548	4564	4579	4594	4609	2	3	5	6	8	9	11	12	14
29	•4624	4639	4654	4669	4683	4698	4713	4728	4742	4757	1	3	4	6	7	9	10	12	13
30	•4771	4786	4800	4814	4829	4843	4857	4871	4886	4900	1	3	4	6	7	9	10	11	13
31	•4914	4928	4942	4955	4969	4983	4997	5011	5024	5038	1	3	4	6	7	8	10	11	12
32	•5051	5065	5079	5092	5105	5119	5132	5145	5159	5172	1	3	4	5	7	8	9	11	12
33	•5185	5198	5211	5224	5237	5250	5263	5276	5289	5302	1	3	4	5	6	8	9	10	12
34	•5315	5328	5340	5353	5366	5378	5391	5403	5416	5428	1	3	4	5	6	8	9	10	11
35	•5441	5453	5465	5478	5490	5502	5514	5527	5539	5551	1	2	4	5	6	7	9	10	11
36	•5563	5575	5587	5599	5611	5623	5635	5647	5658	5670	1	2	4	5	6	7	8	10	11
37	•5682	5694	5705	5717	5729	5740	5752	5763	5775	5786	1	2	3	5	6	7	8	9	10
38	•5798	5809	5821	5832	5843	5855	5866	5877	5888	5899	1	2	3	5	6	7	8	9	10
39	•5911	5922	5933	5944	5955	5966	5977	5988	5999	6010	1	2	3	4	5	7	8	9	10
40	•6021	6031	6042	6053	6064	6075	6085	6096	6107	6117	1	2	3	4	5	6	8	9	10
41	•6128	6138	6149	6160	6170	6180	6191	6201	6212	6222	1	2	3	4	5	6	7	8	9
42	•6232	6243	6253	6263	6274	6284	6294	6304	6314	6325	1	2	3	4	5	6	7	8	9
43	•6335	6345	6355	6365	6375	6385	6395	6405	6415	6425	1	2	3	4	5	6	7	8	9
44	•6435	6444	6454	6464	6474	6484	6493	6503	6513	6522	1	2	3	4	5	6	7	8	9
45	•6532	6542	6551	6561	6571	6580	6590	6599	6609	6618	1	2	3	4	5	6	7	8	9
46	•6628	6637	6646	6656	6665	6675	6684	6693	6702	6712	1	2	3	4	5	6	7	7	8
47	•6721	6730	6739	6749	6758	6767	6776	6785	6794	6803	1	2	3	4	5	5	6	7	8
48	•6812	6821	6830	6839	6848	6857	6866	6875	6884	6893	1	2	3	4	4	5	6	7	8
49	•6902	6911	6920	6928	6937	6946	6955	6964	6972	6981	1	2	3	4	4	5	6	7	8
50	•6990	6998	7007	7016	7024	7033	7042	7050	7059	7067	1	2	3	3	4	5	6	7	8
51	•7076	7084	7093	7101	7110	7118	7126	7135	7143	7152	1	2	3	3	4	5	6	7	8
52	•7160	7168	7177	7185	7193	7202	7210	7218	7226	7235	1	2	2	3	4	5	6	7	7
53	•7243	7251	7259	7267	7275	7284	7292	7300	7308	7316	1	2	2	3	4	5	6	6	7
54	•7324	7332	7340	7348	7356	7364	7372	7380	7388	7396	1	2	2	3	4	5	6	6	7

	0	1	2	3	4	5	6	7	8	9	1	2	3	4	5	6	7	8	9
55	·7404	7412	7419	7427	7435	7443	7451	7459	7466	7474	1	2	2	3	4	5	5	6	7
56	·7482	7490	7497	7505	7513	7520	7528	7536	7543	7551	1	2	2	3	4	5	5	6	7
57	·7559	7566	7574	7582	7589	7597	7604	7612	7619	7627	1	2	2	3	4	5	5	6	7
58	·7634	7642	7649	7657	7664	7672	7679	7686	7694	7701	1	1	2	3	4	4	5	6	7
59	·7709	7716	7723	7731	7738	7745	7752	7760	7767	7774	1	1	2	3	4	4	5	6	7
60	·7782	7789	7796	7803	7810	7818	7825	7832	7839	7846	1	1	2	3	4	4	5	6	6
61	·7853	7860	7868	7875	7882	7889	7896	7903	7910	7917	1	1	2	3	4	4	5	6	6
62	·7924	7931	7938	7945	7952	7959	7966	7973	7980	7987	1	1	2	3	3	4	5	6	6
63	·7993	8000	8007	8014	8021	8028	8035	8041	8048	8055	1	1	2	3	3	4	5	5	6
64	·8062	8069	8075	8082	8089	8096	8102	8109	8116	8122	1	1	2	3	3	4	5	5	6
65	·8129	8136	8142	8149	8156	8162	8169	8176	8182	8189	1	1	2	3	3	4	5	5	6
66	·8195	8202	8209	8215	8222	8228	8235	8241	8248	8254	1	1	2	3	3	4	5	5	6
67	·8261	8267	8274	8280	8287	8293	8299	8306	8312	8319	1	1	2	3	3	4	5	5	6
68	·8325	8331	8338	8344	8351	8357	8363	8370	8376	8382	1	1	2	3	3	4	4	5	6
69	·8388	8395	8401	8407	8414	8420	8426	8432	8439	8445	1	1	2	3	3	4	4	5	6
70	·8451	8457	8463	8470	8476	8482	8488	8494	8500	8506	1	1	2	2	3	4	4	5	6
71	·8513	8519	8525	8531	8537	8543	8549	8555	8561	8567	1	1	2	2	3	4	4	5	5
72	·8573	8579	8585	8591	8597	8603	8609	8615	8621	8627	1	1	2	2	3	4	4	5	5
73	·8633	8639	8645	8651	8657	8663	8669	8675	8681	8686	1	1	2	2	3	4	4	5	5
74	·8692	8698	8704	8710	8716	8722	8727	8733	8739	8745	1	1	2	2	3	4	4	5	5
75	·8751	8756	8762	8768	8774	8779	8785	8791	8797	8802	1	1	2	2	3	3	4	5	5
76	·8808	8814	8820	8825	8831	8837	8842	8848	8854	8859	1	1	2	2	3	3	4	5	5
77	·8865	8871	8876	8882	8887	8893	8899	8904	8910	8915	1	1	2	2	3	3	4	4	5
78	·8921	8927	8932	8938	8943	8949	8954	8960	8965	8971	1	1	2	2	3	3	4	4	5
79	·8976	8982	8987	8993	8998	9004	9009	9015	9020	9025	1	1	2	2	3	3	4	4	5
80	·9031	9036	9042	9047	9053	9058	9063	9069	9074	9079	1	1	2	2	3	3	4	4	5
81	·9085	9090	9096	9101	9106	9112	9117	9122	9128	9133	1	1	2	2	3	3	4	4	5
82	·9138	9143	9149	9154	9159	9165	9170	9175	9180	9186	1	1	2	2	3	3	4	4	5
83	·9191	9196	9201	9206	9212	9217	9222	9227	9232	9238	1	1	2	2	3	3	4	4	5
84	·9243	9248	9253	9258	9263	9269	9274	9279	9284	9289	1	1	2	2	3	3	4	4	5
85	·9294	9299	9304	9309	9315	9320	9325	9330	9335	9340	1	1	2	2	3	3	4	4	5
86	·9345	9350	9355	9360	9365	9370	9375	9380	9385	9390	1	1	2	2	3	3	4	4	5
87	·9395	9400	9405	9410	9415	9420	9425	9430	9435	9440	0	1	1	2	2	3	3	4	4
88	·9445	9450	9455	9460	9465	9469	9474	9479	9484	9489	0	1	1	2	2	3	3	4	4
89	·9494	9499	9504	9509	9513	9518	9523	9528	9533	9538	0	1	1	2	2	3	3	4	4
90	·9542	9547	9552	9557	9562	9566	9571	9576	9581	9586	0	1	1	2	2	3	3	4	4
91	·9590	9595	9600	9605	9609	9614	9619	9624	9628	9633	0	1	1	2	2	3	3	4	4
92	·9638	9643	9647	9652	9657	9661	9666	9671	9675	9680	0	1	1	2	2	3	3	4	4
93	·9685	9689	9694	9699	9703	9708	9713	9717	9722	9727	0	1	1	2	2	3	3	4	4
94	·9731	9736	9741	9745	9750	9754	9759	9763	9768	9773	0	1	1	2	2	3	3	4	4
95	·9777	9782	9786	9791	9795	9800	9805	9809	9814	9818	0	1	1	2	2	3	3	4	4
96	·9823	9827	9832	9836	9841	9845	9850	9854	9859	9863	0	1	1	2	2	3	3	4	4
97	·9868	9872	9877	9881	9886	9890	9894	9899	9903	9908	0	1	1	2	2	3	3	4	4
98	·9912	9917	9921	9926	9930	9934	9939	9943	9948	9952	0	1	1	2	2	3	3	4	4
99	·9956	9961	9965	9969	9974	9978	9983	9987	9991	9996	0	1	1	2	2	3	3	3	4



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